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INCREASING PRODUCTIVITY IN OLD PLANTS THROUGH MODERN MATERIALS HANDLING[1]

BY

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The re-appearance of priorities, allocations and other controls, shortages of materials and manpower, panic buying, and other items in our inventory of revived horrors, serves today to dramatize the fact that there is still a wide abyss between what we have and what we need. American industry is faced again with the compound problem of serving both the prodigious requirements of the country's military organization and the seemingly insatiable demand for consumer goods of all descriptions. During the past few years, fortunately, we have had a respite of comparative peace during which we were able to replace some equipment and to accomplish a considerable amount of new plant construction; but this progress, postponed during World War II, has now been interrupted by the new mobilization of our country.

WHY OLD PLANTS MUST BE BROUGHT UP TO THEIR PRODUCTION PEAKS

The mandatory and voluntary restrictions on materials, together with the panic in the market place, seem to be proof positive that industry is even now hard-put to meet, at the same time, increased civilian demands and the swollen needs of the military. If the "hot-again, cold-again" war has in reality taxed industry's facilities to the full, there remains but one answer: we must either increase productivity or lower our standard of living. In fact, only a few days ago, the Chairman of the President's Committee of Economic Advisors viewed the first alternative—increased productivity—as the only means by which the nation could carry both the civilian and military burden.

[1] Talk presented at the Fifth Annual Meeting of the Society of Industrial Packaging and Materials Handling Engineers, in Convention Hall, Philadelphia, Pa., October 10, 1950.

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(Note—The Franklin Institute is not responsible for the statements and opinions advanced by contributors in the JOURNAL.)

One might say that the obvious course to follow to attain expanded output would be to build new, streamlined plants with ingenious mechanized assembly lines. Unfortunately, not only is there little time, but there's also a growing shortage of building materials and other scarcities which, together with the present shortage of investment capital, preclude the immediate construction of many necessary new plants. Of the open choices, one practical solution, then, is to work with existing plant facilities so as to bring them up to their production peak.

THREE QUARTERS OF ALL PLANTS ARE OVER 25 YEARS OLD

Which brings me to my subject: modern mechanized materials handling methods as a prime tool by which to harvest the fullest capacities of existing plants, old or new.

By recurring demonstrations, it has been shown time and again that handling of all sorts is the "soft underbelly" of production, the sphere of greatest vulnerability in cost and efficiency, the weakest link in the production process. On the other hand, it has been shown, with equal force, that it is in the sphere of handling that industry can make almost incredible progress in reducing cost, in increasing unit output per man hour, and in vitalizing over-all production efficiency. Strengthen this "soft underbelly" with modern handling techniques and you take the long step forward toward increased productivity in existing plants—even the oldest ones.

The recently completed Census of Manufacturers discloses that 72 per cent of all industrial buildings in the United States are over 25 years old. Of these, half are from 35 to 45 years old, another quarter are over 45 years old. This plethora of aging American plants underscores the scope of the problem we are here to discuss today. Old plants and buildings are more likely than not to pose complications which, at first glance, seem to exclude efficient manufacturing. While they may require a very different approach from a new plant, ideally systemized with modern mechanized handling, I believe a combination of engineering ingenuity and imagination, abetted by some modern mechanized materials handling equipment and techniques, can often make a more efficient operating unit of even the oldest plant.

HOW TO DETERMINE "REAL AGE" OF A PLANT

Some old-plant problems include peculiarities and shortcomings in floor strength, low ceiling heights, poor placement of supporting pillars, inadequate aisles and storage bays, and similar disadvantages. But it is not the chronological age of a plant which truly signifies its age, but rather the extent to which these and other disadvantages exist. From a materials handling point of view, and therefore from a production point of view, even a newly-built plant is really an "old plant" if it has these characteristic disadvantages. In a sense, then,

it can be said that a plant has both a chronological age and a technological age, and that all newly-built plants are not necessarily technologically modern.

With this in mind, let's look at how some companies are achieving technological modernity in old plants hitherto plagued by deficiencies in layout and structure.

GEARING RECEIVING AND SHIPPING OPERATIONS FOR MORE PRODUCTIVITY

One of the most crucial points in all materials handling operations is how incoming shipments, whether from highway trucks or freight cars, are treated. Obviously, it is an advantage to have interior docking facilities at just the right height for loading, but these are not usually found in old buildings. Ramps are one solution commonly used where there is a disparity in truck bed height and floor level. If the distance between the highway or rail carrier and the dock is small, simple bridge plates can be used. More expensive, but usually more satisfactory, are hydraulic ramps which can be adjusted to practically any carrier height.

Use of ramps, both inside and outside the plant, however, cause other problems. Since the tractive effort required of lift trucks grows by leaps and bounds with increasing ramp-angles, there is a definite practical grade limitation here (usually 10 per cent).

Where the angle for loading or unloading is initially too great for the use of a ramp, the solution may be excavating, or filling in the receiving bays to bring truck beds nearby even with dock levels. In leased buildings where a minimum of such capital construction is desired, the answer may very well be portable conveyors for unloading.

Some old plants are faced with the problem of loading facilities designed for trucks formerly equipped with side doors. This problem is further complicated because there is often no room for the truck to back up to the loading platform in the usual manner. This has been solved by one company by installing a hydraulic bridge piece that scoots out from under the door of the building at right angles to the tailgate of the truck to make a very satisfactory platform. In addition, when there is no such vehicle to be loaded, the bridge piece is retracted so that the alleyway is clear for through vehicles.

Another common problem in old plants is the lack of headroom and clearance for today's enlarged highway trucks and railroad cars. In addition to receiving doors being too small, many times the receiving area inside the plant is restricted in size because of process operations.

One company solved both of these problems by breaking an opening through the outer wall of the building, about ten feet above ground level, to accommodate a gravity wheel conveyor. Now, incoming shipments are taken from the truck or freight car, placed on the conveyor, and shipped to storage or process areas throughout the plant without any additional loading dock or accessories.

OVERCOMING STRUCTURAL LIMITATIONS OF OLD PLANTS

Once the goods being received are inside, more problems arise in old plants. From a materials handling standpoint, a one-story structure is preferable to a multi-story structure, except for certain process industries. One-story buildings generally afford more usuable space per dollar of investment than any other, partly because of the space requirements of elevators. Furthermore, in a one-story building, relatively new types of roof construction with longer spans, arched and curved welded trusses, designed to support only the roof plus any snow load, offer economies in construction costs.

THE PROBLEM OF ELEVATORS

In multi-story buildings, elevators may also prove to be a bottleneck in the flow of materials. This can partially be avoided by taking into full consideration the volume of material flow and then planning installation of as many elevators with sufficient capacity and door openings as needed to admit and accommodate the loaded trucks.

Automatic systems of signaling and control help reduce truck delays at elevators. The time required for an elevator to respond to a call can be cut by installing the signal switches far enough in front of the trucking lines to enable truck operators to call in advance and by locating the elevator home station at a midpoint in the shaft.

Other ways to prevent elevator delays are to put loads on live skids or dollies which the elevator operator can push on or off or to provide the elevator operator with a motorized hand truck for handling standard skids or pallets. Only the load waits in these two systems.

THE PROBLEM OF FLOOR STRENGTH

Elevators bring up the question of floor strength. There is usually ample floor capacity on ground floors to carry safely almost any type of industrial truck and its load. In extending their operation to upper floors of buildings, the question of adequate floor capacity arises. Since building construction varies widely, it is practically impossible to make recommendations for safe practice without a detailed study of the plant.

An approximate weight of trucks that may be operated safely on a given floor can be indicated by a simple study. Numerous tests made by Yale & Towne's materials handling manufacturing divisions show that industrial trucks impose a dynamic load on the floor of about 25 per cent more than the static weight of the loaded truck. Such test results were obtained in our Materials Handling Testing Laboratories by using strain measurements on the structure and special test setups.

Based on these data, a safety factor of 50 per cent over static load is advised in computations for dynamic floor loading. This means

a safety factor of 1.5. However, it is only one of two factors to consider. An industrial truck exerts its weight for all intents and purposes on a single point. Although the load is actually spread over the wheelbase and the tread of the tires, this is a relatively small area compared to the floor area. This concentrated load produces a bending moment approximately double that produced by a uniform load on the beam structure of the floor. In test, the bending stress actually was only increased 85 per cent so that a safety factor of 2 is more than ample. Combining these two safety factors makes an over-all safety factor of 3:1 for stresses in the floor slab that are imposed by a truck that is in operation.

A SIMPLE METHOD FOR DETERMINING FLOOR LOADING CAPACITIES

The next question is how much floor area is devoted to carrying the weight of the loaded truck. This is usually the aisle area. If the total area of the aisle is multiplied by the capacity of the floor, the total capacity is obtained. Thus a 10-ft. aisle, 20 ft. long with a capacity of 250 psi., would have a total capacity of 10 times 20 times 250, or 50,000 lb. Dividing this number by 3 (the safety factor), the permissible loaded weight of trucks would be 16,000.

If two or more trucks could pass each other over this area, their combined loaded weight could not exceed the total 16,000. This point is very important, *particularly around elevators*, and should be remembered when reinforcing old plants. In addition, the area around elevators usually gets more than its share of traffic, and the fatigue load is likely to be greater. An additional safety factor should therefore be incorporated into the floors near elevators.

These calculations are made for concrete slab floors which, according to architects, are one of the more desirable types for buildings used to support industrial trucks. If other types are used, particularly wood, wheel loading studies may be necessary, since only a few boards or planks may receive the entire load.

If, however, a wooden or similar type of floor will easily support a uniformly distributed load of the magnitude desired, some simple device like steel plates will often serve to disseminate the concentrated load of the wheels over several planks or boards.

When computations indicate that the maximum capacity of the floor is being approached, a competent construction engineer should make a further study, but if there is a generous safety factor, no further effort is usually necessary.

REORGANIZATION OF OLD PLANT FOR BETTER FLOW OF MATERIALS

A good example of how the major objectives of the handling problem can be undertaken by the physical reorganization of an old plant is the project Yale & Towne is undertaking at its Stamford Division.

This program is designed to lower the break-even point in our manufacture of Yale locks and builders' finishing hardware.

In the Stamford plant we are confronted with the consequences of venerability. The first building was erected in 1868, and since then, additions were made until 1928. The plant is now a series of six-story buildings, linked together in the form of a huge "L" sprawled over a 29-acre site. Inside this "L" of buildings, on a center quadrangle, are miscellaneous buildings for metal processing, woodworking, power, shipping, and receiving. Production has in the past consisted of specialized operations for each product regardless of where the product had to travel—from building to building, or from shop to shop—in the process.

The Stamford General Manager is eliminating this method of manufacture and in the new procedure each of the six floors will have a continuous production system based on these classifications of product: Raw material brought into these floors—and in this case, it consists of material upon which primary operations have been completed in foundries and press shops—will go through a one-floor cycle of secondary operations. Each floor will have four similar areas—an area for secondary machining operations, a storage area for machined parts, an assembly area, and a packaging area.

Production is planned to take place on a straight line system. Parts will be mechanically handled on belt and hanger conveyors, and monorail hoist systems as well as mobile equipment such as hand trucks, motorized hand trucks, and lift trucks.

Over-all, the Stamford Division of Yale & Towne is consolidating its present 92 departments scattered over the entire plant into 44 locations in the "L" of buildings. Each of the six floors, in effect, will be a single-story unified operation for a group of related products. The present indications are that a movement economy of about 50 per cent will be achieved.

HOW TO ACHIEVE MORE EFFICIENT HANDLING IN SEPARATED BUILDINGS

Where a plant must use several buildings of the multi-story variety, additional handling headaches are often introduced. Too often, in the past, these buildings were not connected, and material sometimes had to proceed from the fifth floor of one building to the fifth floor of the next via an arduous trip down an elevator across a roadway and up another elevator. There have been some interesting methods devised to correct this evil. Here are a few:

Floors between the buildings can be connected with bridges which serve as truck runways. This will eliminate the tedious elevator trip and will speed handling of materials from building to building as well as leave the elevators free for other work. If there is bulk material to

be handled, a bulk conveyor, of the screw, vacuum, or blower type, can be used. For handling cartons, pallets, and other uniform loads, a tray type or gravity roller conveyor has been found successful. Often maximum economy is achieved by complementing a truck bridge and runway with a conveyor.

An ingenious idea for connecting two multi-story buildings was developed by a large paper box company. This manufacturer had two buildings that were separated by a gap of ten feet, although for reasons of "unifying" the property, the outer walls had been connected. Practically all inter-floor transportation was inter-building transportation. Material went down in one elevator and up in the other. There were two small elevators in one of the buildings, a six-story affair which served as the factory, and an old elevator in the six-story adjacent warehouse. The two elevator shafts in the factory were located in the center of the building, taking up valuable production space, and were not wholly adequate to handle the flow of materials from floor to floor when it became necessary to transfer so much material from one building to the other.

The management realized the prohibitive cost of this antiquated operation and arranged for an elevator company to make a survey. The engineers recommended the placing of an elevator shaft between the two buildings for the installation of a modern freight elevator and also a passenger elevator. A great improvement in the efficient flow of materials resulted when the plan was put into effect and the old elevators removed.

When installing the new elevators, the difference in the elevation between floor levels in the two adjacent plants was a special problem. This was solved by using elevators which have doors on two sides. The new elevator takes its load on the fourth floor of the manufacturing building, then goes up two feet and discharges the material in the warehouse building for distribution to stock.

Where old one-story plants are separated by a wide open area, it has often been found that roofing the area over will not only improve efficiency of operations, but also provide additional warehouse space. It usually provides a more useful area too. Although the yard may have been used for storage previously, covered with a roof, the area becomes more liable to orderliness and good housekeeping.

TECHNOLOGICAL METHODS, NOT AGE OF PLANT, DETERMINE ITS MODERNITY

As I have been talking here, you have probably thought of many other solutions to problems of mechanized handling in old plants. Each plant will have certain problems and no two will be alike even in plants making the same product in the same type of building.

Two factors that most frequently govern the solution of modernizing materials handling methods in old plants, or for that matter any

production modernization, are the setup that now exists, and the amount that you can spend to replace the present system with something better. Other important factors you might add to these are the amount of engineering imagination and ingenuity available, either in your own organization or in the consulting engineers, and how much free rein it is given.

And it is important to remember that no old plant is absolutely hopeless. Certainly a plant, old or new, may lack flexibility, and may not please the eye in appearance, but it is not the brick and mortar shell that governs the unit cost of its products, it is how you fabricate them, how you assemble them, and how you *handle* them. In conclusion, gentlemen, it is not the age of the building but the technological methods employed that really determine the modernity of any plant.

THE GOODNESS OF FIT AND THE DISTRIBUTION OF PARTICLE SIZES [1]

BY

F. KOTTLER [2]

Part I [3]

SUMMARY

In a previous paper (1),[4] reasons were given why the usual graphical analysis of the distribution of particle sizes should be replaced by an algebraic analysis. In the present paper, this method is developed and explained. Most of the mathematical argument has been removed from the text and is given in several appendices. The method of analysis described is comparatively new in that it uses the Chi-Square Minimum Principle instead of the more familiar principle of least squares.

By the results of this analysis, it is possible to characterize the size distribution in a photographic emulsion, in the simpler cases, by two numbers (parameters): The first, denoted by a, which is related to the time of growth of the crystals, and the second, denoted by b, which is inversely proportional to the velocity constant of the growth. The first parameter is subject to statistical fluctuations within an individual emulsion; the second should be constant for one type of emulsion as long as the physicochemical conditions of growth are not altered.

This theory is confirmed by careful analysis of one sample of an emulsion, eight subsamples of which were measured by the observer. The results of the statistical analysis show that the parameter, b, is the same for the eight subsamples within the limits of the experimental error. The values of the parameter, a, show, however, differences in excess of those limits. This is due to the local fluctuations of the time of growth throughout the emulsion, mentioned above. Statistically, there is, therefore, *heterogeneity* in one of the *population* parameters. Therefore, the results from the eight subsamples should not be added together. This is usually done with the intention to increase the goodness of the fit, but actually decreases it.

INTRODUCTION

In the previous paper (1), it was pointed out that the sizes of particles (silver halides, sand, dust, pigments, etc.) are log-normally distributed. Usually this distribution is analyzed by means of log-probability graphs only. It was pointed out that this kind of analysis is inaccurate and, in many cases, leads to spurious results. The log-normal distribution must rather be analyzed by *algebraic* methods.

Moreover, the mere analysis or, more accurately, *estimation* of the parameters of the distribution is not sufficient. To the estimation must be added a *test* of the goodness of the fit, which is only possible with *algebraic* analysis.

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[1] Communication No. 1338 from the Kodak Research Laboratories, Eastman Kodak Company, Rochester, N. Y.

[2] Kodak Park Works, Research Laboratories, Eastman Kodak Company, Rochester, N. Y.

[3] Part II will appear in this JOURNAL for June 1951.

[4] The boldface numbers in parentheses refer to the references appended to the second part of this paper.

In the algebraic analysis, the graphs still play an important role. If treated carefully, these graphs furnish a rough first approximation, the parameters of which can be read off the graph. These parameters are then corrected algebraically by a series of further approximations. These corrections are small if the graphical analysis has been made carefully, which is possible for a well-trained observer. For an inexperienced observer, the corrections of the values obtained by the graphical analysis will, as a rule, be very large and will necessitate a long series of approximations.

Another advantage of the graphs is that they can be used to illustrate each of the consecutive steps of the analysis and so provide a better insight into the algebraic process.

THE FREQUENCY POLYGON

As is well known, there are two curves used in problems of this kind, the cumulative distribution curve (ogive) and the non-cumulative (bell-shaped) frequency curve. Of these, the former is widely used in graphical analysis, chiefly because it is represented by a straight line on a log-probability chart. It is the only curve used in Probit Analysis and in psychophysics. However, it is not the best basis for the algebraic analysis of the present problem.

This is seen as follows: The ogive represents the probabilities,

$$P = F/N,$$

where F is the cumulative frequency reached at the end-point[5] of an interval, for example,

$$F_1 = f_1,\ F_2 = f_1 + f_2,\ F_3 = f_1 + f_2 + f_3 \cdots.$$

Here, f_1 is the class-frequency of the class-interval 1, f_2 is that of class 2, etc. ($N = f_1 + f_2 + \cdots + f_n$).

Now, suppose there is an error, δ_1, in the observed class-frequency, f_1, an error, δ_2, in f_2, etc. Obviously there will be an error, $\delta_1 + \delta_2$, in the cumulative frequency, F_2, an error, $\delta_1 + \delta_2 + \delta_3$, in F_3, etc. That is to say, any F is affected not only by the error of one class but also by the errors of all the preceding classes. Hence, the F's are not independent of one another but are mutually correlated. To avoid misunderstanding, we repeat that this is not the case in Probit Analysis nor in psychophysics. For instance, if the response of insects to poison is tested as in Probit Analysis (1), the divers F's are taken from different samples and so are independent of each other.

To avoid the aforementioned difficulty, we must use the class-frequencies, f, instead of the cumulative frequencies, F. It is true that

[5] This applies to an "ascending" ogive, such as we shall use here exclusively. There is also a "descending" ogive, used, for instance, in sieve analysis of powders, where

$$F_1 = N,\ F_2 = N - f_1,\ F_3 = N - f_1 - f_2, \text{ etc.},$$

where the cumulation starts from the upper end instead of the lower end.

even the class-frequencies, f, and their errors, δ, are not quite independent of each other. However, for the time being, we shall regard them as completely independent of one another, because that simplifies matters and permits the application of the familiar method of least squares. Later, we shall take into consideration the slight interdependency of the f and of the δ and shall replace, consequently, the method of least squares by the less familiar one of Chi-Square Minimum.

There is a certain difficulty in the graphical presentation of the class-frequencies. Unlike the cumulative frequencies, F, the class-frequencies, f, are not *point-functions*. As a matter of fact, they are *interval-functions* and depend not upon *one* but upon *two* points, the beginning and the end of the class-interval. The class-frequency, therefore, depends not only upon a location, but also upon a length, the width of an interval. The latter dependence can be eliminated by "standardizing" the class-frequency, that is, by referring it to unit interval. This is done by considering the ratio,

$$f/\Delta x,$$

where Δx is the width of the interval. The location, on the other hand, is arbitrarily chosen as the *mid-point* of the interval, so it becomes possible to plot the above quantities, $f/\Delta x$, on a graph *versus* the mid-points. By joining the points so obtained, we have a *frequency polygon* (see Fig. 1, solid line).

The frequency polygon is similar to, but not identical with, the *frequency curve*. The latter is obtained by differentiating the cumulative frequency function, F, with respect to x:

$$\frac{dF}{dx},$$

whereas the former is obtained by differencing:

$$\frac{\Delta F}{\Delta x}.$$

Here, the operator, Δ, denotes the difference between two consecutive frequencies:

$$\Delta F_2 = F_2 - F_1 = f_2, \text{ etc.}$$

As is well known from calculus, the two quotients, the differential and the difference, are not the same; the former is the limit of the latter only when the width of the interval decreases indefinitely. Consequently, the frequency curve and the frequency polygon differ from each other not inconsiderably. This can be seen from Fig. 1, in which the data for Seed Lantern No. 6102, Plate III, have been used. (See later Table I.) The frequency curve is represented here by a broken line and has been calculated from the results of the final estimation (see Table V,

columns 1–5). The frequency polygon, on the other hand, is given directly by the observed class-frequencies and the chosen class-intervals.

The difference between frequency curves and frequency polygons is especially noticeable in Class 1; see points P' on the frequency curve and P on the frequency polygon. This is of purely mathematical origin and has nothing to do with any physical causes which have sometimes been quoted for the explanation of that discrepancy. (See Appendix I.)

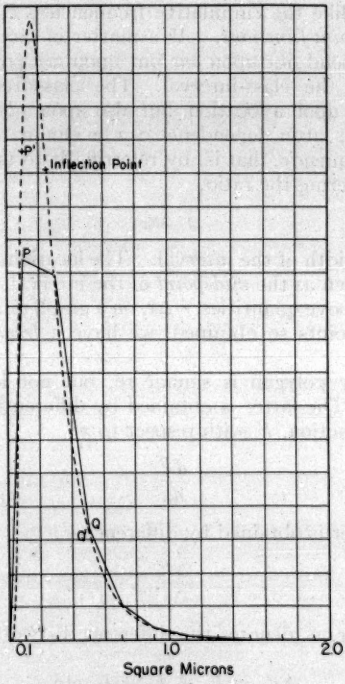

FIG. 1. Frequency polygon and frequency curve. (Seed Lantern No. 6102, III.)

From the above argument, it can be seen easily that the frequency polygon is in practice more important than the frequency curve. The latter cannot be obtained directly from observation. It is, of course, possible to obtain the cumulative frequency curve directly from observation, and it would seem a simple task to derive the frequency curve from the cumulative frequency curve by the process of differentiation. However, this differentiation would have to be a graphical one and as such is

known to be wholly unreliable. Moreover, it has already been mentioned that the cumulative frequency curve is objectionable because it cumulates the errors. Therefore, it is itself an unreliable source from which to derive, indirectly, another curve.

Furthermore, there is another argument in favor of the frequency polygon as a basis of the analysis. As has been mentioned, the process of *estimation* of the parameters of the distribution must be followed by the process of *testing*. The testing quite obviously must be done directly with the data and not with numbers derived indirectly, by some complicated process, from the data. We shall, therefore, use here exclusively the frequency polygon as a basis of the estimation and the testing. The frequency curve will be omitted from the following developments for the reasons given. It is true that the frequency curve is the most popular of the three curves, ogive, non-cumulative frequency curve, and frequency polygon. It is also true that it is independent of the choice of the intervals as is the ogive. However, the choice of the intervals is as much an integral part of the experimental evidence as the observations of the class-frequencies and once chosen must be accepted as an unalterable basis of the testing. (See later the theory of the Chi-Square test.)

The choice of the class-interval is usually made so that it is constant for each class. However, there are cases where we have to deal with different intervals for different classes. It goes without saying that the shape of the frequency polygon depends upon the width of the interval and varies, for example, if the observations are pooled for several classes or are subdivided among subclasses of a single class.

LOGARITHMIC CHARTS

Loveland and Trivelli (2) introduced a graphical presentation of the log-normal frequency curve (broken line in Fig. 1) by means of a *logarithmic* chart[6] (see Fig. 2). This chart is divided according to a logarithmic scale on each axis. The log-normal frequency curve (broken line) appears as a parabola on this chart. This is easily seen from Appendix I of the previous paper: The log-normal frequency curve is given by the expression,

$$1/\sqrt{2\pi}\exp\{-1/2(a + 1/b + b \ln x)^2\}.$$

Its logarithm is, therefore, a quadratic function of $\ln x$. Plotting that logarithm *versus* $\ln x$ gives a parabola which is vertical and opens downward. It has its vertex at a point, the coordinates of which are the logarithms of the mode and of $1/\sqrt{2\pi}$, respectively. Its focus lies at a distance of $b^2/8$ beneath that vertex. The quantity b is the slope of the

[6] It should not be called "doubly logarithmic," since this suggests the iterated logarithm and is, moreover, not the commercial name. The chart in Fig. 1 is, in contrast, called an *arithmetic* one.

log-normal ogive if represented as a straight line on a log-probability chart; also, $1/b$ is a measure of dispersion and is proportional to the velocity constant of growth, k.

In Fig. 2 there is also plotted the frequency polygon by a full line. This, for the reasons given before, is the only one we shall use in the future. It will be seen that there is a great discrepancy at the mid-point of Class 1 between the frequency curve (point P') and the frequency polygon (point P). (See Appendix I of the present paper.) At the remaining points, the lines do not differ much.

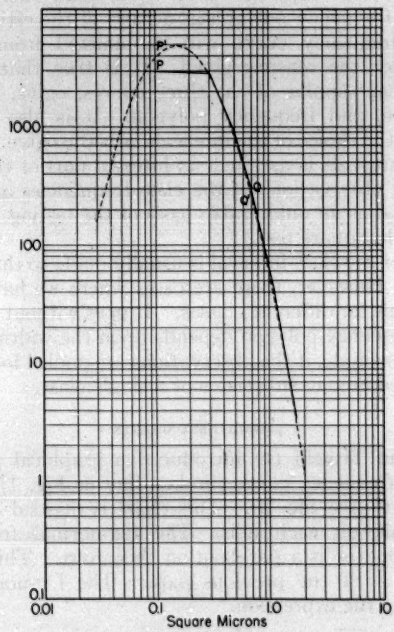

FIG. 2. Frequency polygon and frequency curve. (Seed Lantern No. 6102, III.)

Obviously, a graphical analysis by means of a logarithmic chart is still more difficult than that by means of a log-probability chart which was discussed in the previous paper (1). The parabola on the logarithmic chart does not lend itself as easily to graphical treatment as does the straight line on the log-probability chart. Commercially available templates of parabolas do not cover all cases. Construction with the aid of a compass and the theory of conic sections would demand that the focus and the parameter of the parabola be known, which is not the case.

Moreover, the logarithmic chart distorts the true dimensions as does the log-probability chart. It cannot, therefore, be used to judge the goodness of the fit of any parabola to given data. For instance, in Fig. 2 the vertex of the parabola has the ordinate, 5000, and the lower right end is near the ordinate, 1. An absolute deviation of, say, 10, will be, therefore, quite inconspicuous at the vertex, since the ordinates, $5000 \pm 10 = 5010$ or 4990, are indistinguishable from 5000. At the lower right end, the same absolute error would cover a full cycle. It is true that the relative error is correctly represented by a logarithmic chart. For instance, an error of one per cent appears equal at the vertex of the parabola and at the lower right end. However, the errors are not to be judged by their absolute nor by their relative values but by their standardized values obtained after division by the standard error (as has been explained in the previous paper (1)).

Because of the reason just mentioned, we shall use error bands (1). Since the standard error depends upon the total number, N, of grains counted, the frequency polygon *must not be reduced* to $N = 1$ or 1000, as is sometimes done. Consequently, the expression given before for the log-normal frequency curve must be multiplied by N so that the area beneath it is equal to N instead of being equal to unity (or 1000).[7] As pointed out earlier (1), only the error bands render a chart usable for an approximate judgment on the goodness of the fit. Without these bands, a chart is completely useless for such a purpose.

In what follows we shall use the logarithmic charts as an illustration of the algebraic process only and shall not attach any importance to the parabolic character of the frequency curve on such charts. We shall, moreover, use exclusively the frequency polygon which is not a parabola. The chief advantage of the use of the logarithmic chart is that the right tail of the curve appears on it (Fig. 2) greatly magnified as compared with its appearance on an arithmetic chart (Fig. 1). It goes without saying that this magnification is accompanied by a great deal of distortion; the reader should, therefore, be warned not to attach any importance to the absolute dimensions as they appear on a logarithmic chart near the right tail.

THE PRINCIPLE OF THE METHOD OF LEAST SQUARES

We now turn to the algebraic treatment of the problem and will begin with a few remarks, which may not be familiar to many readers, on the method of least squares.

We shall regard that method as derived from a principle of maximum probability. This is not always done but is here necessary for several reasons: First, the method should be linked with a law of errors which has been well established. This is at variance with the attitude assumed

[7] The frequency polygon is always *unreduced*, since by definition it is given by the values $f/\Delta x$ and $\Sigma f = N$.

by Gauss, himself, in his later work (15). However, it should not be forgotten that we deal here with observations which are governed by some law which can be established only by experience, as is the law of errors. Next, the presentation of the method in the form of a so-called "maximum principle" has many advantages when mathematical proofs are required. Finally, this presentation allows an easy comparison between the method of least squares and the Chi-Square Minimum method which will replace the former in later sections of this paper.

We assume that the errors, $\delta_1, \delta_2, \cdots \delta_n$, in the observed class-frequencies, $f_1, f_2, \cdots f_n$, each obey the Normal Law of Errors, that is, the probability of an error, δ_1, is proportional to

$$\exp\left(-\delta_1^2/2s_1^2\right),$$

where s_1 is the standard error for Class 1. Similarly, for the other classes. Then, assuming that all the n errors are independent of each other, the joint probability of the simultaneous occurrence of them will be given by the product of these n probabilities, which is proportional to

$$\exp\left(-\delta_1^2/2s_1^2\right)\exp\left(-\delta_2^2/2s_2^2\right)\cdots\exp\left(-\delta_n^2/2s_n^2\right) = \exp\left\{-1/2\sum\delta^2/s^2\right\},$$

where the summation extends over the n classes. This joint probability becomes a maximum when the sum,

$$\sum \delta^2/s^2,$$

becomes a minimum. This is the principle of the method of least squares.

SUCCESSIVE APPROXIMATIONS

In general, the problem of minimizing the aforementioned sum cannot be solved in one step. The solution in one step is only possible in the familiar problem of linear regression. In other cases, one would be led to Normal equations which are not linear in the unknowns. Such equations would, therefore, furnish more than one system of solutions and one would have to discuss each of them before rejecting those solutions which in one way or another do not suit the physical conditions of the problem. The problem has to by *linearized* so as to get linear Normal equations. This is done as follows: Let

$$y = f(x, a, b)$$

be the equation to be fitted to the data, y. Here, a, b, are the unknown values of the parameters which we want to find by the method of least squares. The data, y, will, of course, not fit exactly but, owing to the errors, will leave "residuals," δ, so that

$$y = f(x, a, b) + \delta.$$

Let it be assumed that the values, a', b', are a first approximation for the unknown values, a, b. Then, we may use a Taylor series for the function, f, with respect to the increments, α', β', which are defined by

$$a = a' + \alpha', \quad b = b' + \beta'.$$

We obtain

$$f(x, a, b) = f' + \alpha' \frac{\partial f'}{\partial a'} + \beta' \frac{\partial f'}{\partial b'} + \cdots,$$

where f' signifies the values obtained from the first approximation. The Taylor series has been terminated after the terms of the first degree, assuming that the increments, α', β', are so small as to permit that. (Actually, this is usually but not always permissible. We shall return to this point later.)

We may, therefore, write

$$y = f' + \alpha' \frac{\partial f'}{\partial a'} + \beta' \frac{\partial f'}{\partial b'} + \delta,$$

and also,

$$\delta = \delta' - \alpha' \frac{\partial f'}{\partial a'} - \beta' \frac{\partial f'}{\partial b'},$$

where $\delta' = y - f'$ are the residuals left by the first approximation.

We introduce now the *weights* which are defined as being inversely proportional to the squares of the standard errors, that is,

$$w = \frac{\epsilon^2}{s^2}, \tag{1a}$$

where the factor, ϵ^2, is left indeterminate. Then we have to minimize the sum,

$$\sum w\delta^2,$$

with respect to the increments, α', β'. Substituting for δ and performing the partial differentiations with respect to α', β', we obtain:

$$\sum w \left(\delta' - \alpha' \frac{\partial f'}{\partial a'} - \beta' \frac{\partial f'}{\partial b'} \right) \frac{\partial f'}{\partial a'} = 0,$$

$$\sum w \left(\delta' - \alpha' \frac{\partial f'}{\partial a'} - \beta' \frac{\partial f'}{\partial b'} \right) \frac{\partial f'}{\partial b'} = 0.$$

These are the Normal equations for the unknowns, α', β'. They are linear, as they should be, and therefore furnish a unique set of solutions. We shall write them in the form:

$$\begin{aligned} a_{11}\alpha' + a_{12}\beta' &= a_{13}, \\ a_{21}\alpha' + a_{22}\beta' &= a_{23}, \end{aligned} \tag{1b}$$

where we have set:

$$a_{11} = \sum w \frac{\partial f'}{\partial a'} \frac{\partial f'}{\partial a'} \quad a_{12} = \sum w \frac{\partial f'}{\partial b'} \frac{\partial f'}{\partial a'} \quad a_{13} = \sum w \delta' \frac{\partial f'}{\partial a'}.$$

$$a_{21} = \sum w \frac{\partial f'}{\partial a'} \frac{\partial f'}{\partial b'} \quad a_{22} = \sum w \frac{\partial f'}{\partial b'} \frac{\partial f'}{\partial b'} \quad a_{23} = \sum w \delta' \frac{\partial f'}{\partial b'}.$$

Note the symmetry, $a_{12} = a_{21}$, etc. The summation is to be extended over all n classes of observations. The matrix,

$$\begin{Vmatrix} a_{11} & a_{12} & a_{13} \\ a_{21} & a_{22} & a_{23} \end{Vmatrix},$$

will be called the Normal matrix. The determinant,

$$\begin{vmatrix} a_{11} & a_{12} \\ a_{21} & a_{22} \end{vmatrix},$$

will be called the Normal determinant. Its minors will be denoted as follows:

$$A_{11} = a_{22}, \qquad A_{12} = -a_{12} = -a_{21}, \qquad A_{22} = a_{11}.$$

The inverse matrix to the Normal determinant will be denoted by

$$\alpha_{11} = A_{11}/A, \qquad \alpha_{12} = A_{12}/A, \qquad \alpha_{22} = A_{22}/A,$$

where A signifies the value of the Normal determinant. Note that the Normal determinant and its inverse are symmetric and also that they are independent of the data, y. Only the right-hand sides, a_{13}, a_{23}, do depend upon the data.

In the practical computation of the Normal matrix, it is customary to compute the elements of the Normal matrix as follows:

$$\begin{matrix} [A'A'] & [A'B'] & [A'E'] \\ [B'A'] & [B'B'] & [B'E'], \end{matrix}$$

where the brackets are the Gaussian symbols for the summation, Σ, over all classes and

$$A' = \sqrt{w} \frac{\partial f'}{\partial a'}, \qquad B' = \sqrt{w} \frac{\partial f'}{\partial b'}, \qquad E' = \sqrt{w}\, \delta'.$$

In this way, the symmetry of the matrix is more clearly revealed; this symmetry affords an important check in the practical computation of the elements. So far we have discussed only one algebraic approximation, the second approximation, based on a first non-algebraic one, the parameters of which have been chosen in one way or another. (In our case, it will be derived, usually, from a graphical approximation on log-probability paper.) It may happen that the second approximation is not sufficient. (Criteria of sufficiency for this operation will be given

below.) In that case we have to repeat the process and to proceed as follows: Let

$$a'' = a' + \alpha', \qquad b'' = b' + \beta'$$

signify the values obtained by the second approximation. Let us start a third approximation,

$$a = a'' + \alpha'', \qquad b = b'' + \beta'',$$

where the new increments, α'', β'', are to be determined by the method of least squares. The Taylor series used in that method will, of course, now start with the values, f'', obtained from the second approximation.

To shorten the procedure, it is obvious that a good first approximation ought to be chosen so as to make a third, fourth, etc., unnecessary and to render the second approximation sufficient and final. In order to judge the sufficiency of any approximation, we must know the standard errors of the solutions, α, β, derived from it.

THE STANDARD ERRORS OF THE SOLUTIONS

The solutions, α', β', of the Normal equations (1*b*) are obtained in the usual way indicated by Gauss. First, one unknown is eliminated, say, α', by expressing it as a linear function of the other unknown, β'. There results a single equation for that unknown whose solution is straightforward. In the same way, a system of more than two Normal equations and more than two unknowns can be solved by successive eliminations.

The solutions so obtained are linear functions of the data, $y_1, y_2 \cdots y_n$. They are, therefore, subject to the experimental errors of the data. Gauss has shown how to compute the standard errors of the solutions, α', β', from the standard errors of the data. (See Appendix II.)

Gauss's formula is as follows:

$$s_{\alpha'}{}^2 = \alpha_{11}\,\epsilon^2, \qquad s_{\beta'}{}^2 = \alpha_{22}\,\epsilon^2. \tag{2}$$

The factor, ϵ^2, is determined as follows:

$$\epsilon^2 = \sum w\delta'^2/(n-2),$$

or (3)

$$\epsilon^2 = \sum w\delta''^2/(n-2) \text{ etc.},$$

according to the approximation under consideration. This formula is not exact, since the sum in the right-hand expression refers to the actual series of observations only and ought to be replaced by its mean value after the observations are repeated many times. That mean value is, however, not known generally and, therefore, Eq. 3 is generally used.

TERMINATING THE SUCCESSIVE APPROXIMATIONS

The material given in the last three sections is due to Gauss and is therefore contained in most of the textbooks on the treatment of observations. We come now to several points which go beyond that.

This is not quite true for the first point to be discussed, namely: when should the chain of successive approximations be terminated? Gauss, himself, touched on this question lightly (see a quotation from Gauss in Deming's book (4)), but he did not offer a general rule for dealing with that question.

We state here such a rule without proof: An approximation shall be considered final and no further approximation shall be necessary, if the corrections, α, β, do not exceed their respective standard errors, s_α, s_β. On the other hand, if one or both of them do exceed these limits, a further approximation shall be necessary. In most cases, it will be found that the second approximation is already sufficient, provided the first approximation was a good start.

Incidentally, we shall sometimes have to compare different sets of solutions, α, β, arrived at by different ways or from different data. They shall be regarded as equal to each other within the limits of the experimental error, if they differ by less than *twice* their respective standard errors. If the standard errors of the two sets are not equal, we shall take the *larger* values of the standard errors as a basis for that comparison.

THE CHI-SQUARE DISTRIBUTION

Another point which goes beyond the achievements of Gauss and his followers during the nineteenth century concerns the judgment or *test of the goodness of the fit* obtained by *estimating* the unknown parameters by means of the method of least squares. This is effected by the distribution of the Chi-Square, that is, of the sum,

$$\chi^2 = \sum \delta^2/s^2, \tag{4}$$

over the squared and standardized residuals. This sum was known to Gauss and his followers as a rough estimate of the goodness of the fit, but no exact rules were established. The first to introduce the statistical distribution of the Chi-Square was Helmert (5), although his paper seems not to have met with general acceptance. The so-called Chi-Square test came into general use only after Karl Pearson (6) introduced it into statistics.

According to Pearson, the goodness of the fit is judged by considering the value of the Chi-Square instead of the individual errors, δ. This is as it should be, since it is not possible to make all individual data fit the computed curve. To consider individual points and their fit as it is still practised, especially in the purely graphical methods, means, under these circumstances, to open the door to arbitrariness and to bias. Instead, one should consider the over-all fit of all the data together and not of several arbitrarily selected ones. In choosing the Chi-Square for such an over-all criterion, some sacrifices must be *accepted*. The Chi-Square deals only with the squared errors, as does the method of least squares, and therefore neglects the signs of the errors. Whether

these signs are all equal or whether they alternate at random is the same to the Chi-Square test. This is a deficiency of that test which cannot be overcome. No other test which would take account of the signs as well as of the absolute magnitudes of the errors has, as yet, been generally accepted.[8]

Obviously, the Chi-Square should be as small as possible, if the fit is a good one. A critical limit is, therefore, defined which the Chi-Square must not exceed, if the fit is to be regarded as acceptable. This limit is set arbitrarily (from the results of practical experience) as follows: There is a certain probability for each numerical value of the Chi-Square which can be computed from the probabilities for the individual errors, *provided each of them obeys the Normal law of errors*. Consider then the probability that the Chi-Square does exceed the critical value. This probability is stipulated to be 10 per cent or more. If the Chi-Square exceeds the critical value, the observations are said to *deviate significantly* from the hypothesis underlying the computation and that hypothesis ought to be *rejected*, in the opinion of many statisticians. As already mentioned, we disagree, since we substitute a well-established *theory* for the *hypothesis* of the statisticians; hence, we shall, in such a case, doubt the observations rather than the theory and shall repeat them several times before rejecting the theory.

The level of 10 per cent is called the rejection level. There are other levels in use, as 5 per cent, or 1 per cent. They impose less stringent conditions on the goodness of the fit but will not be used in the present paper. To each of these levels corresponds a different critical value of the Chi-Square. It can be found from tables of the Chi-Square distribution as they are given in many textbooks (see Cramér (7), Kendall (8), Fisher and Yates (9)).

THE INTERDEPENDENCY OF THE ERRORS

The method of least squares described in the previous sections does not apply to our problem, because of the interdependency of the errors, which is contrary to the assumption stated in the previous section on the principle of the method of least squares.

That the errors are interdependent can be seen as follows: The data in our case, obey the equation,

$$\Sigma f = N,$$

in which f signifies the observed class-frequencies. Hence, the computed class-frequencies, f' (or f'', etc.), must also obey such an equation,

$$\Sigma f' = N,$$

since the total number of grains counted, N, is a primary datum of the experiment which cannot be changed. (Later we shall see how that

[8] Another deficiency of the Chi-Square test will been countered later in dealing with the small class-frequencies.

condition is enforced in the computations.) From these two equations we have, therefore, the fundamental equation,

$$\sum \delta = 0. \tag{5}$$

Hence, the errors influence each other. For instance, if $n - 1$ errors are given, the nth error is no longer arbitrary but completely determined by them.

Therefore, the method of least squares does not apply. It will be replaced by the method of Chi-Square Minimum. Fortunately, both methods are formally similar to each other, so that most of what has gone before can be retained and only minor variations are necessary (Appendix III).

THE PRINCIPLE OF CHI-SQUARE MINIMUM

As shown in Appendix III, we have, for the probability of the simultaneous occurrence of n errors, δ, which are bound by Eq. 5, the following approximate expression:

$$\text{Const. exp } (- \chi^2/2).$$

Here we have set:

$$\chi^2 = \sum \delta^2/\sigma^2 \tag{6}$$

in which, as explained in Appendix III,

$$\sigma = \sqrt{f'}. \tag{7}$$

This replaces the former Eq. 4, used in the method of least squares.

Formally, we are again faced with a problem of maximum probability. We may inquire about the values of the parameters, a, b, which render the above exponential a maximum, and hence, its exponent a minimum. This suggests a method which is called the method of Chi-Square Minimum and which can deal with interdependent errors that the method of least squares could not handle.

It is interesting to compare the present method with the method of Probit Analysis and of psychophysics already discussed (1), where the method of least squares was applied to the ogive. This was legitimate, since the errors in probability, δP, were independent of each other, each one referring to a different sample. We used the expression,

$$\chi^2 = \sum (\delta P)^2/S^2$$

with $S^2 = F'(1 - F'/N)$ and F' denoting the cumulative frequencies.

Now, we are studying the class-frequencies, f', and are using

$$\chi^2 = \sum \delta^2/\sigma^2$$

with $\sigma^2 = f'$ instead of $s^2 = f'(1 - f'/N)$, which would be analogous to the above formula. Moreover, the number of degrees of freedom was $n - 2$ in Probit Analysis, whereas it is now $n - 3$, because of the interdependency of the errors due to Eq. 5.

In spite of this difference, our present Chi-Square obeys the same distribution as the former, as was shown for a similar case by R. A. Fisher (10). One need only take care of the difference in the degrees of freedom.[9]

We shall call the weights, $1/\sigma^2 = 1/f'$, used in Eq. 6, the *Pearson weights*, to distinguish them from the Urban weights, $1/S^2 = 1/F'(1 - F'/N)$ used in the previous paper (1).[10] Karl Pearson (6, pp. 160–164) was the first to consider the Chi-Square of a frequency distribution.

THE MODIFIED CHI-SQUARE MINIMUM METHOD

The problem of minimizing the form given by Eq. 6 is a non-linear one, for several reasons: First, the numerator is non-linear in the unknowns, a, b, as has already been explained in dealing with the similar form of the method of least squares. Moreover, the denominator contributes, now, to the non-linearity, because it depends upon the unknowns, a, b (Eq. 7). In the older form (4), the mean square errors do not depend upon the unknowns.

Recourse to successive approximations takes care of both non-linearities. Assume that a first approximation, f', with the approximate values, a', b', be given. We then attempt to find a second approximation by determining the increments, α', β', in

$$a' + \alpha', \qquad b' + \beta',$$

so that the new Chi-Square shall be a minimum. As before, we shall assume a Taylor expansion, the higher terms of which we neglect. We obtain

$$\sum \delta''^2/f'',$$

which must be a minimum. Here

$$f'' = f' + \alpha' \frac{\partial f'}{\partial a'} + \beta' \frac{\partial f'}{\partial b'} + \cdots,$$

$$\delta'' = \delta' - \alpha' \frac{\partial f'}{\partial a'} - \beta' \frac{\partial f'}{\partial b'} - \cdots,$$

and $\delta' = f - f'$ is the residual for the first approximation, $\delta'' = f - f''$ is that for the second. We can neglect the terms with α', β' in the denominator[11] and write

$$\sum \delta''^2/f'$$

[9] "Fundamentally. . .these are merely different aspects of the same thing and rest on the fact that the variation of the sum of squares of normal variates in standard measure is spherically symmetric, so that a hyperplane in the sample space 'cuts' the distribution in a spherically symmetric form of one lower degree of freedom." Kendall, Ref. 8, Vol. 2, p. 158.

[10] Or $1/s^2 - 1/f'(1 - f'/N)$, which would be analogous to the Urban weight in our present case.

[11] They add a term of the third order, if δ' is assumed to be small of the same order as α', β'.

for the expression to be minimized. This expression has been called the *modified* Chi-Square (see Cramér 7, p. 426). It uses weights which are independent of the unknowns, α', β', and depend only upon the approximate values of the parameters, a', b'. *In our presentation, this modification follows naturally from the introduction of successive approximations or what amounts to the same thing from the necessity of linearizing the minimum problem in order to get a unique set of solutions.*[12]

It must, however, be borne in mind that the modified Chi-Square cannot be used in testing the goodness of the resulting fit. For that purpose, we must always use the *unmodified* Chi-Square.

The computation by the modified Chi-Square Minimum proceeds, in the main, formally along similar lines to the ordinary method of least squares. The main difference between the two methods is that the residuals of the Chi-Square Minimum show a peculiar behavior which is due to their interdependency by Eq. 5. An example of the computation will be found in a succeeding section.

POOLING THE SMALL CLASS-FREQUENCIES

As explained in Appendix III, the Chi-Square distribution holds only if the class-frequencies are not small. For those classes which have small frequencies, it is not permitted to presume a Normal distribution of their errors, which is necessary for the Chi-Square distribution.

In this case, it has become customary to pool several classes with small frequencies so that a single class with a larger frequency results. (The class-interval for the pooled class is, of course, the sum of the individual class-intervals.) The rule adopted for defining small class-frequencies has already been mentioned and is as follows: Every computed class-frequency should be at least 10. Classes with frequencies smaller than 10 must be pooled so as to give one or several classes, each with a frequency equal to or exceeding 10. (See Cramér, 7, p. 420.)

This refers only to the application of the Chi-Square test *after* the completion of the computation. Pooling should not be applied to the computation, that is, the process of *estimation* of the parameters. For instance, the method of least squares takes care of the irregular behavior of the small class-frequencies automatically, and, similarly, the method of Chi-Square Minimum does so, as may be stated here without proof. Pooling before computation would only lose degrees of freedom which would render the results more uncertain. Moreover, it would introduce a certain amount of bias by the manipulation of the data in pooling.

(*To be continued.*)

[12] The method of the modified Chi-Square gives results equivalent to those obtained by the method of *maximum likelihood* introduced by R. A. Fisher and much used in statistics. We have chosen here the former method because it is based on the familiar concept of probability rather than on the unfamiliar concept of likelihood.

THORIA AS A CATHODE EMITTER[1]

BY

W. E. DANFORTH[2]

ABSTRACT

This paper outlines the cathode characteristics of thoria which lead to its selection in certain cases where the barium-strontium oxide cathode is unsuitable. Dependable values of pulsed thermionic emission of 4 amps/cm.^{2} at 1600° C. and 14 or 15 amps/cm.^{2} at 1800° C. are available. The decay of emission with time is much less rapid than with barium-strontium oxides, the time constants being of the order of one-tenth of 1 second. Maximum d-c. emission, that is, equilibrium emission after decay, is found to be between 1 and 5 amps/cm.^{2}. Although when compared to barium-strontium oxide the activation phenomena are not prominent, it is found that the same essential pattern exists. For use in pulsed applications the life of a thoria coated cathode is determined by its evaporation rate. At 1800° C. where 14 amps/cm.^{2} is available the coating disappears at the rate of 1 mil thickness in about 300 hours.

INTRODUCTION

Among the technological developments during the recent war was the use of thoria for electron tube cathodes in situations where the usual barium-strontium oxide coatings could not withstand the punishment to which they were subjected. Operating at a higher temperature, the thoria cathodes naturally required more heater power. In this article are outlined the characteristics of thoria which justify the additional cathode heat. These may be placed in two groups: (1) *Thoria is more refractory* than barium-strontium oxide. Greater resistance to arcing, sparking, etc. is largely due to the relatively high melting point and lower rate of evaporation. (2) *The activation of thoria is less easily disturbed* than that of barium-strontium oxides. The initial processing is less troublesome. The activated cathode is relatively immune to poisoning and can be operated in a poor vacuum where barium oxide would quickly deactivate. The decay of emission with time, at high currents, is much less rapid than with barium oxide, permitting longer pulse widths. Larger d-c. values are available although the life with high d.c. is not all that is desired.

1. THERMIONIC EMISSION

We turn now to the thermionic emission of thoria as found by several observers. We mean by this the so-called pulsed emission, the value of emission before appreciable decay has occurred.

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[1] This was an invited paper delivered at the Symposium on Thermionics of the American Physical Society, January 21, 1950. The author is indebted for much of the material here presented to other members of the thermionics group at the Bartol Foundation: D. L. Goldwater, E. M. Shapiro and O. A. Weinreich.

[2] Assistant Director, Bartol Research Foundation of The Franklin Institute, Swarthmore, Pa.

In Fig. 1 we have plotted data from several sources. The solid curve is drawn through those data which are regarded as representative of stable pulsed emission. Dependable values of 4 amps/cm.^{2} at 1600° C. true temperature and 14 or 15 amps/cm.^{2} at 1800° C. are obtained.

In addition to this stable emission, however, a thoria cathode can be further activated to a higher level. This enhanced emission shown on the graph by the upper dashed curve and solid circles is of unstable character and, unless special means are resorted to, will deactivate in relatively short time to the level of the solid curve.

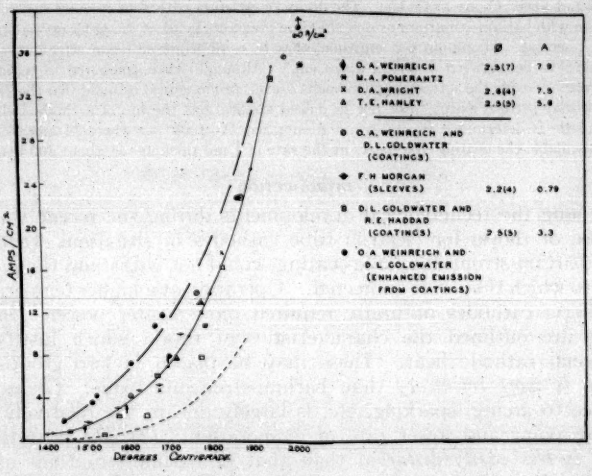

FIG. 1.

Values of Richardson constants as obtained by several observers are given on Fig. 1.

The data shown by the open squares were taken in a bell jar under poor vacuum conditions. We believe them to be of interest, however, as evidence of the relatively high stability of a thoria emitter in the pressure region between 10^{-5} and 10^{-4} mm. The data were repeatable with several different specimens and were independent of pressure in the region cited.

The behavior of thoria in poor vacuum has been reported in some detail by Wright.[3] He finds, among other things, that the sensitivity of a thoria cathode to oxygen is much less than that of a thoriated tungsten cathode, even when the latter is carbonized.

[3] D. A. WRIGHT, *Proc. Phys. Soc.*, Vol. B62, p. 188 (1949).

2. ACTIVATION PHENOMENA

When one who is accustomed to BaSr oxide cathodes turns his attention to the activation of a thoria cathode his first reaction is likely to be that no activation is required. Indeed, if a thoria cathode is heated to 1700° or 1800° for such length of time as outgassing may require its emission will probably be fully activated to the level given by the solid curve in Fig. 1. This behavior at first suggests that one is dealing with an intrinsic emitter rather than an impurity activation.

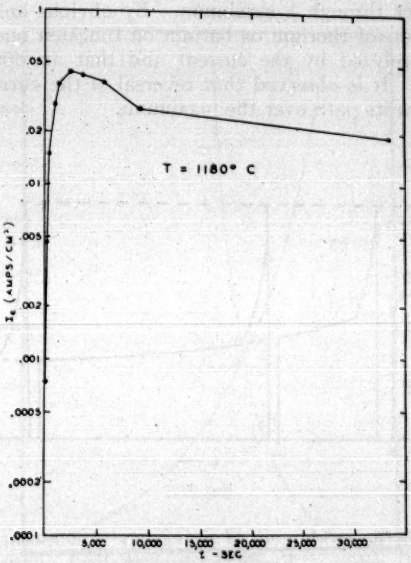

FIG. 2.

However, it has been found by Fan,[4] and independently at the Bartol Foundation,[5] that the behavior of thoria at low temperatures (between 1200° and 1400°) indicates a complex activation pattern.

Activating processes which have been studied are of three kinds: (1) flashing to high temperature, (2) subjection to flow of current, especially in the reverse direction, and (3) exposure to vapor from a nearby source.

Process (*1*) (flashing) could provide impurity activation either by

[4] H. Y. FAN, *J. Appl. Phys.*, Vol. 20, p. 682 (1949).

[5] O. A. WEINREICH, *Phys. Rev.*, Vol. 75, p. 1312 (A) (1949); *J. Appl. Phys.*, Vol. 20, p. 1256 (L) (1949).

generating free thorium by reduction, or by evaporation or diffusion of metallic tungsten in the thorium oxide lattice. The free tungsten hypothesis is not entirely ruled out but, for several reasons, the free thorium hypothesis appears more likely if not, indeed, fairly well established. The presence of free thorium in an activated cathode can be convincingly demonstrated by its evaporation upon, and thermionic activation of, a nearby tungsten wire.

Process (*2*)—activation by flow of current in the reverse direction—is shown in Fig. 2. One sees that, as time progresses, the emission rises and passes through a maximum. By obvious analogy with the behavior of films of thorium or barium on tungsten one assumes that thoria is electrolyzed by the current and that an optimum surface density exists. It is observed that reversal of the current causes the curve to retrace its path over the maximum.

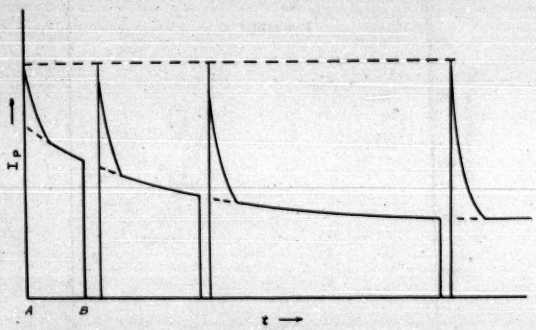

FIG. 3.

Process (*3*) is the attainment of activation by evaporation of material from an outside source. From the hypothetical picture it would be expected that evaporation of thorium metal would produce activation. As yet, to our knowledge, this has not succeeded. It has been found possible, however, to activate by evaporation products from another thoria cathode. It is found that a somewhat complex behavior exists. The donator cathode may give off either poisoning material or activating material depending on the temperature where it is used.

Although the free thorium hypothesis accounts satisfactorily for the broad features of the situation, some outstanding questions remain to be answered, such as, for example, the failure to secure activation by evaporated thorium metal. Also D. A. Wright[3] points out that it is difficult to account quantitatively for the rapidity of activation on the free thorium hypothesis.

3. REVERSIBLE DECAY

The d-c. emission of a thoria cathode should not be regarded as a specific property of thorium oxide but is rather the equilibrium emission which a cathode can maintain in the presence of certain deactivating circumstances. Prominent among these possibilities are: (1) electrolytic generation of oxygen at the surface, (2) diffusion of thorium away from the surface when the field is applied, (3) release of poisoning material from the bombarded anode, (4) destruction of a surface film by ion bombardment.

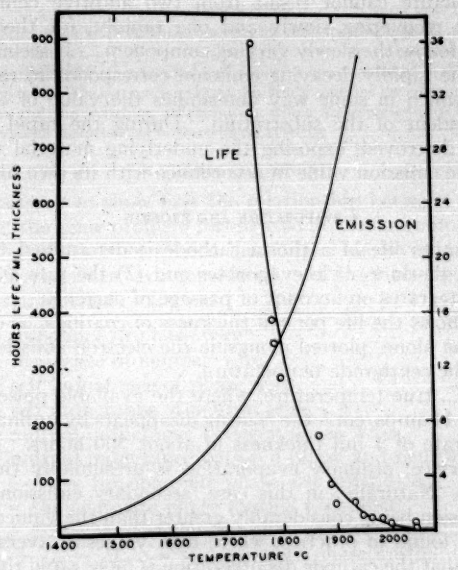

FIG. 4.

The phenomena of reversible decay are not present with pure metals but have great prominence with BaSr oxide. Both as regards magnitude of decay and its rapidity, thorium oxide has an intermediate position. For example, while a BaSr oxide cathode will show a large decay component with a time constant of 10^{-5} seconds, the most rapidly decaying component with thoria has a time constant of a few tenths of one second.

The occurrence of decay limits the available d-c. emission to values which do not exceed 5 amps/cm.^{2} and may in many cases be less than 1 amp/cm.^{2}.

Figure 3 is an idealized diagram which illustrates a feature of decay phenomena which has bearing on the general character of the mechanism involved. In this experiment the anode voltage is applied, for example, at point A, and the emission permitted to decay to point B where the voltage is removed and recovery takes place. The time of recovery permitted is not sufficient to allow the slowly varying component to recover. However, it is found that under these circumstances the emission will recover its initial value almost completely but will then decay rapidly.

Such a picture cannot result from two additive components of emission, one recovering slowly and one rapidly, for then the peak values would follow the slowly varying component. It seems to require rather that the rapidly decaying emission corresponds to removal of a surface film which in some way determines the value of emission by itself, independent of the substratum. During the rapid decay this film is being destroyed, exposing the underlying material which then determines the emission value in accordance with its own history.

4. EVAPORATION AND EROSION

The maximum life of a thoria cathode is determined by: (1) the rate at which thorium oxide evaporates and (2) the rate at which the material disintegrates on account of passage of current.

Figure 4 shows the life per mil thickness of coatings, as determined by evaporation alone, plotted alongside the electron emission, both as functions of the centigrade temperature.

At 1800° C. true temperature, where the available pulsed emission from thoria is 14 amps/cm.^{2}, the coating disappears by ordinary evaporation at the rate of 1 mil thickness in about 300 hours. For pulsed magnetron service, ordinary evaporation is presumably the essential factor in life. Naturally, in this case, secondary emission results in the total emission being considerably greater than the values on Fig. 4.

It is often found in practice, where high values of average current are involved, that the cathode disintegration is more rapid than straight evaporation would account for. For example, in certain tests at Bartol, where sintered cathodes were operated for 1000 hours at 200 ma/cm.^{2}, it was seen that at 1700° C. the sleeves eroded at the rate of 1 mil thickness in about 100 hours. Such a rate of disappearance could be explained by electrolysis if one assumes that ionic conduction in the sleeve accounts for approximately one part in ten thousand of the total current.

Observation of the disintegrative effects of current, however, are far from consistent and further investigations are necessary to explain the variations.

TRANSIENT ANALYSIS OF THREE-PHASE POWER SYSTEMS, PART II *

BY

LEONARD RABINS [1] AND ERIC T. B. GROSS [2]

SUMMARY

The difficulties encountered with symmetrical components, as discussed in Part I, B, are eliminated when Clarke components are used. It is possible to investigate transient phenomena in three-phase systems by studying single-phase networks with the transient analyzer. The basic relationships between the three-phase problem and its equivalent representation in Clarke components are established in the first section (C) of this paper. The last section (D) of this study contains the component-network connections and the switching arrangements which are representative of the more frequent types of disturbances in three-phase systems.

C. CLARKE COMPONENTS

It is necessary to show first the relationship between a three-phase problem and the same problem formulated in Clarke component quantities. A three-phase transmission line (Fig. 1) is investigated in which only series impedances[3] are considered. It shall be assumed that all phases are symmetrical. Let

R_a = the resistance of phase wire,
R_n = the resistance of neutral[4] wire,
L_a** = the self inductance of phase wire,
L_n = the self inductance of neutral wire,
L_{ab} = the mutual inductance between phase wires,
L_{an} = the mutual inductance between phase and neutral wires,
$R_{11} = R_a + R_n$ = the complete phase self resistance,
$R_{12} = R_n$ = the complete phase mutual resistance,
$L_{11} = L_a + L_n - 2L_{an}$ = the complete phase self inductance, and
$L_{12} = L_{ab} + L_n - 2L_{an}$ = the complete phase mutual inductance.

* This paper is based on a thesis supervised by Eric T. B. Gross and submitted by Leonard Rabins to the Faculty of the Graduate School, Illinois Institute of Technology, in partial fulfillment of the requrements for the degree of M. S. in Electr. Eng. Part I appeared in this JOURNAL, March 1951, p. 333.

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[3] In transient analysis, the term impedance applies to either operational impedance or to the operator $\left(R + L\frac{d}{dt} + S\int_{-\infty}^{t} dt\right)$.

[4] The terms "neutral" and "ground" shall be used interchangeably.

** Here inductances will not be defined in terms of circuit configurations. Our interest lies with sequence inductances which are easily found from basic flux and current relationships. See reference 27.

Generated voltages shall be designated by e, terminal voltages by v, and currents by i. Instantaneous quantities shall be designated by lower case letters with corresponding subscripts for further identification.

The mesh equations for the three-phase networks are in matrix form[5]

$$\begin{bmatrix} e_a \\ e_b \\ e_c \end{bmatrix} = \begin{bmatrix} R_{11} R_{12} R_{12} \\ R_{12} R_{11} R_{12} \\ R_{12} R_{12} R_{11} \end{bmatrix} \begin{bmatrix} i_a \\ i_b \\ i_c \end{bmatrix} + \begin{bmatrix} L_{11} L_{12} L_{12} \\ L_{12} L_{11} L_{12} \\ L_{12} L_{12} L_{11} \end{bmatrix} \cdot \begin{bmatrix} \frac{d}{dt} i_a \\ \frac{d}{dt} i_b \\ \frac{d}{dt} i_c \end{bmatrix} + \begin{bmatrix} v_a \\ v_b \\ v_c \end{bmatrix}, \quad (14)$$

or

$$[E] = [R][I] + [L]\frac{d}{dt}[I] + [V]. \quad (15)$$

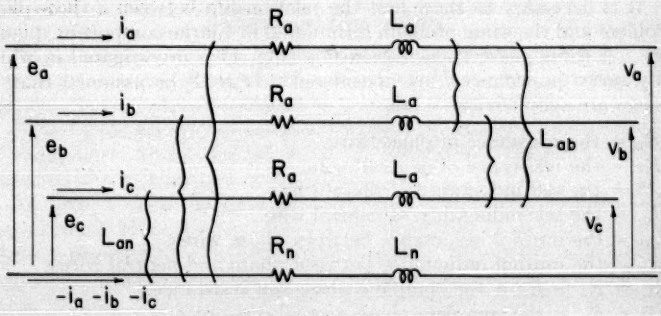


FIG. 1. Three-phase circuit showing series impedances.

It will be necessary to define the Clarke component quantities. For this purpose we shall call the transformation matrix

$$[C] = \begin{bmatrix} 1 & 1 & 0 \\ 1 & -\frac{1}{2} & \frac{1}{2}\sqrt{3} \\ 1 & -\frac{1}{2} & -\frac{1}{2}\sqrt{3} \end{bmatrix}. \quad (16)$$

The inverse matrix is then

$$[C]^{-1} = \frac{1}{3}\begin{bmatrix} 1 & 1 & 1 \\ 2 & -1 & 1 \\ 0 & \sqrt{3} & -\sqrt{3} \end{bmatrix}. \quad (17)$$

[5] For an elementary treatment of matrices, see reference 28.

The various component quantities are defined as follows for currents:

$$[I_s] = \begin{bmatrix} i_0 \\ i_\alpha \\ i_\beta \end{bmatrix} = [C]^{-1}[I] = \tfrac{1}{3}\begin{bmatrix} i_a + i_b + i_c \\ 2i_a - i_b - i_c \\ 0 + \sqrt{3}i_b - \sqrt{3}i_c \end{bmatrix}; \quad (18a)$$

and voltages:

$$[E_s] = \begin{bmatrix} e_0 \\ e_\alpha \\ e_\beta \end{bmatrix} = [C]^{-1}[E] = \tfrac{1}{3}\begin{bmatrix} e_a + e_b + e_c \\ 2e_a - e_b - e_c \\ 0 + \sqrt{3}e_b - \sqrt{3}e_c \end{bmatrix}. \quad (18b)$$

From the definition of the inverse matrix, it follows that

$$[I] = \begin{bmatrix} i_a \\ i_b \\ i_c \end{bmatrix} = [C][I_s] = \begin{bmatrix} i_0 + i \\ i_0 - i_a/2 + \sqrt{3}i_\beta/2 \\ i_0 - i_a/2 - \sqrt{3}i_\beta/2 \end{bmatrix}. \quad (19)$$

A similar relation applies to voltages. Transforming Eq. 15*b* to component quantities, we have

$$[C]^{-1}[E] = [C]^{-1}[R][C][C]^{-1}[I] + [C]^{-1}[L][C][C]^{-1}\frac{dt}{dt}[I] + [C]^{-1}[V]. \quad (20)$$

The transformation has the property of diagonalizing a symmetrical matrix. For example,

$$[C]^{-1}[R][C] = \begin{bmatrix} R_{11} + 2R_{12} & 0 & 0 \\ 0 & R_{11} - R_{12} & 0 \\ 0 & 0 & R_{11} - R_{12} \end{bmatrix} = \begin{bmatrix} R_0 & 0 & 0 \\ 0 & R_1 & 0 \\ 0 & 0 & R_1 \end{bmatrix}. \quad (21)$$

The values of component impedances are designated with the subscripts "0" and "1", since these impedances are identical with the zero-sequence and positive-sequence impedances obtained by transformation to symmetrical components. A similar relationship and definition applies to inductances. Writing out the result in full

$$\begin{bmatrix} e_0 \\ e_\alpha \\ e_\beta \end{bmatrix} = \begin{bmatrix} R_0 & 0 & 0 \\ 0 & R_1 & 0 \\ 0 & 0 & R_1 \end{bmatrix}\begin{bmatrix} i_0 \\ i_\alpha \\ i_\beta \end{bmatrix} + \begin{bmatrix} L_0 & 0 & 0 \\ 0 & L_1 & 0 \\ 0 & 0 & L_1 \end{bmatrix}\begin{bmatrix} \frac{d}{dt}i_0 \\ \frac{d}{dt}i_\alpha \\ \frac{d}{dt}i_\beta \end{bmatrix} + \begin{bmatrix} v_0 \\ v_\alpha \\ v_\beta \end{bmatrix}, \quad (22a)$$

or

$$[E_s] = [R_s][I_s] + [L_s]\frac{d}{dt}[I_s] + [V_s]. \qquad (22b)$$

The subscript s emphasizes the fact that these matrices are the component quantities. This equation indicates the power of the transformation. The variables are separated from one another and all quantities are expressed in terms of quantities of components only. A physical interpretation of the equation shows that individual component networks may be set up (Fig. 2). This set of three single-phase equivalent circuits replaces the original three-phase circuit in the analysis. This equivalent can be extended to a large three-phase circuit, having both series and shunt elements. It shall be shown that each circuit element

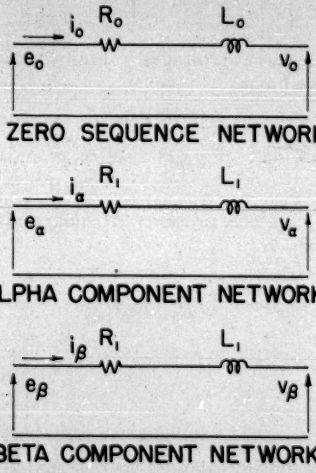


FIG. 2. Clarke component networks; equivalent for Fig. 1.

has a unique representation in the component networks. Various relationships between component quantities at the faulted point[6] lead to definite connections between the component networks at the point of fault. Thus, the three-phase problem is replaced by a single-phase problem in which no mutual impedance terms are present.

We shall now consider the transformation of shunt elements and specifically capacitive elements. The three-phase circuit is shown in Fig. 3. The capacitances[7] of the circuit can be represented as existing

[6] Fault shall indicate anything which upsets the balance of the system.

[7] Here capacitance will not be defined in terms of circuit configurations. Our interest lies with sequence capacitances which are easily found from basic electric field and charge relationships. See reference 27.

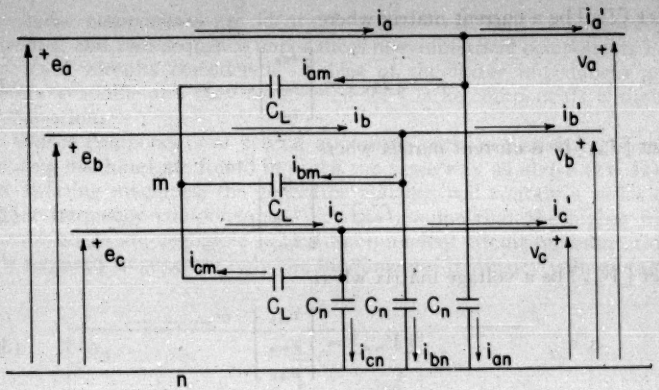


FIG. 3. Three-phase circuit showing shunt capacitances.

between phase wires and between the phases and neutral. The former capacitances constitute a delta set and can be replaced by an equivalent wye set as shown in the figure. Four sets of equations are required for a complete analysis.

$$[E] = [V] \tag{23a}$$

and transforming

$$[E_s] = [V_s]. \tag{23b}$$

Now let $[S_n]$ be an elastance[8] matrix where

$$[S_n] = \begin{bmatrix} S_n & 0 & 0 \\ 0 & S_n & 0 \\ 0 & 0 & S_n \end{bmatrix}. \tag{24}$$

Let $[S_L]$ be an elastance matrix where

$$[S_L] = \begin{bmatrix} S_L & 0 & 0 \\ 0 & S_L & 0 \\ 0 & 0 & S_L \end{bmatrix}. \tag{25}$$

Let $[I_n]$ be a current matrix where

$$[I_n] = \begin{bmatrix} i_{an} \\ i_{bn} \\ i_{cn} \end{bmatrix}. \tag{26}$$

Let $[I_m]$ be a current matrix where

$$[I_m] = \begin{bmatrix} i_{am} \\ i_{bm} \\ i_{cm} \end{bmatrix}. \tag{27}$$

[8] Elastance is the reciprocal of capacitance.

Let $[I_{ns}]$ be a current matrix where

$$[I_{ns}] = \begin{bmatrix} i_{0n} \\ i_{\alpha n} \\ i_{\beta n} \end{bmatrix}. \tag{28}$$

Let $[I_{ms}]$ be a current matrix where

$$[I_{ms}] = \begin{bmatrix} 0 \\ i_{\alpha m} \\ i_{\beta m} \end{bmatrix}. \tag{29}$$

Let $[V_{mn}]$ be a voltage matrix where

$$[V_{mn}] = \begin{bmatrix} v_{mn} \\ v_{mn} \\ v_{mn} \end{bmatrix}. \tag{30}$$

Let $[V_{mns}]$ be a voltage matrix where

$$[V_{mns}] = \begin{bmatrix} 3v_{mn} \\ 0 \\ 0 \end{bmatrix}. \tag{31}$$

Then

$$[E] = [S_n] \int_{-\infty}^{t} [I_n]\, dt \tag{32a}$$

and transforming

$$[E_s] = [S_n] \int_{-\infty}^{t} [I_{ns}]\, dt. \tag{32b}$$

Also,

$$[E] = [S_L] \int_{-\infty}^{t} [I_m] dt + [V_{mn}] \tag{33a}$$

and transforming

$$[E_s] = [S_L] \int_{-\infty}^{t} [I_{ms}] dt + [V_{mns}]. \tag{33b}$$

Also,

$$[I] = [I_n] + [I_m] + [I'] \tag{34a}$$

and transforming

$$[I_s] = [I_{ns}] + [I_{ms}] + [I_s']. \tag{34b}$$

These equations lead to an equivalent circuit shown in Fig. 4. This analysis for shunt capacitances applies in exactly the same manner to shunt loads and line conductances.

Thus it is seen that in transforming the three-phase circuits to a set of component networks each element, current, voltage, and impedance retains its identity. This fact immediately allows the extension to complete circuits in which series and shunt elements occur together. It is worthwhile to notice once more that the alpha, beta, and zero impedances

of Clarke components are identical, respectively, with the positive, negative, and zero sequence impedances of symmetrical components for the static circuits considered. Tables of the latter impedances are readily available and may be adopted for Clarke components without modification.

Clarke components of special apparatus such as transformers and rotating machines are found in much the same way as above (29, 32).[9] For rotating machines, the generator voltages will contain a series of higher frequency components. We may assume that the higher frequency terms are negligible so that in equivalent circuit representation it is required to generate only the fundamental frequency voltage (35).

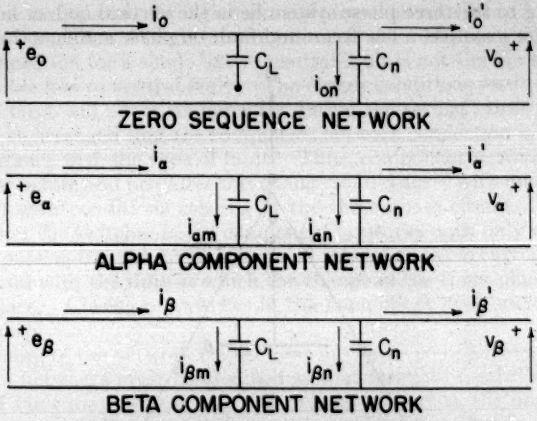


FIG. 4. Clarke component networks; equivalent for Fig. 3.

One note should be added regarding impedances to be used in the equivalent circuits. These impedances are determined for low frequency or direct current conditions. Certainly they are not exact even for sixty-cycle conditions and definitely not during transient conditions. However, the variations are small and it is best to assume that the impedances are constant. We can only hope that the assumption of constant parameters for lines, transformers, and machines gives us solutions within the range of engineering accuracy. Tests bear out this hope (31). The specific study of equipment with non-linear characteristics such as rotating machines, transformers and lightning arresters often must be done using three-phase equivalent circuits. This paper will not deal with the analysis when non-linear impedances are involved although for some simple types of faults no changes in the methods are necessary.

[9] The boldface numbers in parentheses refer to the references appended to both parts of this paper.

It has been shown so far how the various elements of a three-phase system are represented in its equivalent networks of Clarke components.[10] Considerations of the relationships between the component quantities at the fault terminals lead to a set of equivalent circuits of the component networks. This work has previously been done (2, 32) and it has been shown here that these circuits may be adopted for transient studies.

To illustrate the physical meaning of Clarke components consider Fig. 5. The triangle *abc* represents the line voltages of a three-phase balanced system. Alpha and zero component voltages lie in the vertical axis while the beta component voltages lie in the horizontal axis. All phasors rotate at the system frequency. All changes or unbalances occurring to the three-phase system lie in the vertical and/or horizontal axis. For example, a line to ground fault on phase *a* affects the voltage

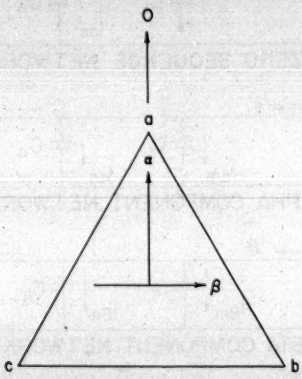


FIG. 5. Phasor relationships between a three-phase circuit and Clarke component circuits.

from *a* to *n*. This change occurs in the vertical axis so that only the alpha and zero networks are affected. A similar fault on phase *b* causes a change which may be resolved into both axes so that all three networks are affected. The network connections for these types of faults substantiate this reasoning. Continuing, a line to line fault occurring on phases *b* and *c* produces changes only in the horizontal axis so that only the beta network is affected. A double line to ground fault causes changes in both axes so that all networks are affected.

D. EQUIVALENT CIRCUITS FOR FAULT REPRESENTATIONS

To extend the method of Clarke components to transient analysis by means of a network calculator is the definite purpose of this investigation. Each important type of fault will be considered separately and if

[10] A derivation of the long line equations in Clarke components appears in the Appendix.

it is necessary to investigate others these methods can easily be applied. Several general observations are common to all cases and these will be noted first.

No difficulty is encountered in determining the three-phase quantities from the component quantities since the transformation matrices involve only real numbers. Thus no modification of the oscillograms of the component quantities is necessary at all. Phase-shifting transformers are not required; again because transformations involve only real quantities.

Regarding changing faults, consider a system having a double line to neutral fault. A definite connection among the component networks exists. Then suppose that one of the phase currents in the fault conductor (or arc) passes through zero and does not restrike. After this time, the circuit has a single line to ground fault. The component network connection for a single line to neutral fault is not the same as that for a double line to neutral fault. The circuit conditions resulting from the first fault will affect its behavior during the second; thus switches must be so arranged that the component network connection is changed in accordance with the type of fault. Thus, component network conditions just before and just after this change should agree with one another and also with conditions existing in the three-phase circuit. This requires that the switches in the component networks open only when the current passing through them is zero and this time of zero current should correspond with the time at which the change in the three-phase circuit takes place. Closing of switches in the component networks presents no problem.

In many of the network connections in Clarke components, the zero and beta networks must be modified so that terminal conditions at the point of fault may be realized. The change occurs in the impedances and generated voltages of the networks. Consider a balanced three-phase system suddenly undergoing a single line to ground fault on phase b. Before the fault the networks are not interconnected and none is modified. To satisfy the fault conditions the three networks are put in series but now the zero sequence impedances are doubled, beta sequence impedances are tripled, beta generated voltages are increased by $\sqrt{3}$, and accordingly beta currents are decreased by $\sqrt{3}$. The conditions before the fault will affect the behavior after and these are realized through the circuits' initial conditions at the instant of fault. Now considering the beta network, before the fault currents corresponding to i_β were flowing in the network. Immediately after the fault these currents must correspond to $i_\beta/\sqrt{3}$ and all impedances are increased by three. The question arises as to whether the initial conditions corresponding to i_β and Z_β shall be effective after the fault or conditions corresponding to $i_\beta/\sqrt{3}$ and $3Z_\beta$. The answer is furnished simply by using Laplace Transform analysis. This tells us that initial conditions due to inductance are equal to a generated voltage of magnitude Li^0, and those due to

a capacitance are equal to a voltage of magnitude q^0/C, where i^0 and q^0 are the current and charge initial values at the instant of change. Since all beta component voltages in this equivalent circuit are increased by $\sqrt{3}$, all initial condition effects which are a form of generated voltage must also be increased by this factor. Let $e_\beta{}^0$ represent an initial condition in the beta network.
Then

$$e_\beta{}^0 = i_\beta{}^0 L_\beta \quad \text{or} \quad q_\beta{}^0/C_\beta \tag{35}$$

and with the required modifications

$$\sqrt{3}\, e_\beta{}^0 = \frac{i_\beta{}^0}{\sqrt{3}}\, 3L_\beta = \sqrt{3}\, i_\beta{}^0 L_\beta \quad \text{or} \quad \frac{q_\beta{}^0}{\sqrt{3}}\,\frac{3}{C_\beta} = \frac{\sqrt{3} q_\beta{}^0}{C_\beta} \tag{36}$$

and this result is exactly what is desired. Since these changes in circuit voltages, currents, and impedances cannot take place instantly, the circuit must be modified before the change is made. Any measurements in the modified circuit must be considered in terms of the modification. This modification has no effect on the circuit behavior except with respect to the magnitude of the various voltages and currents. The

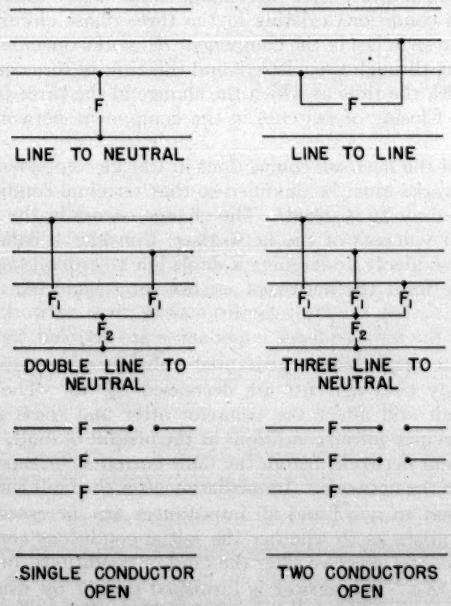
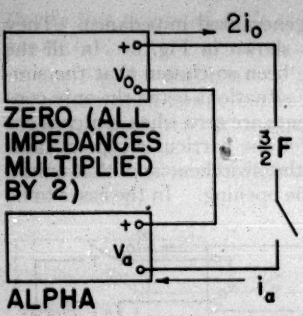


FIG. 6. Location of fault impedances.



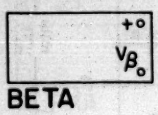
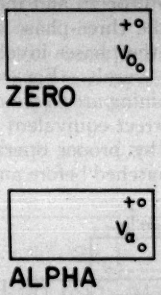


FIG. 7.



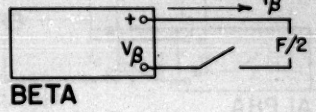


FIG. 8.

Equivalent circuits for fault representations.

following equations show this point. Let the equation for any mesh in the circuit be

$$L\frac{di}{dt} + Ri + \frac{1}{C}\int_{-\infty}^{t} i\,dt = e. \tag{37}$$

Suppose all currents are reduced by $\sqrt{3}$, all voltages increased by $\sqrt{3}$, and all impedances (circuit parameters) increased by three. For this we have an equation

$$3L\frac{di/\sqrt{3}}{dt} + 3R\frac{i}{\sqrt{3}} + \frac{3}{C}\int_{-\infty}^{t}\frac{i}{\sqrt{3}}\,dt = \sqrt{3}\,e. \tag{38}$$

All the multipliers cancel and the same equation as above is obtained. Thus the transient frequencies, decrement factors, etc., remain the same. Only the magnitudes of the terms desired change.

The modification of circuit parameters may be achieved by an actual change of the physical values or by the use of an ideal[11] transformer with a proper turns ratio located at the fault terminals. The only restriction is that once the component networks are connected, then when a change occurs and the new condition requires that these component networks again be connected no further modification shall be needed.

In the following, the component networks will be identified as a box with the fault terminals emanating from it. Fault impedances will be

[11] The transformer should be so ideal that it should have no appreciable effect on the transient behavior of the system.

denoted by F, F_1, or F_2 and indicate a generalized impedance. They are located in the three-phase circuit as shown in Fig. 6. In all the following cases, the phases involved have been so chosen that the simplest connections result. For each of the situations listed the only considerations remaining are to see that currents are zero when switches are opened; that correct equivalent circuits for the particular type of fault are maintained by proper operation of the switches; and that initial conditions are matched before and after the opening. In the cases to be

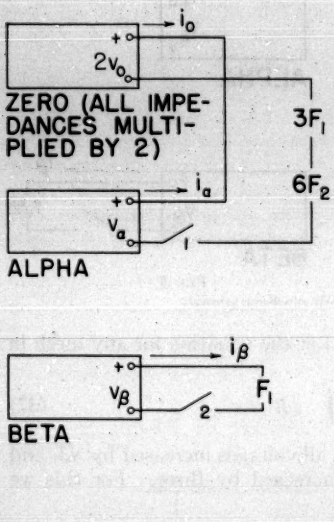


FIG. 9.

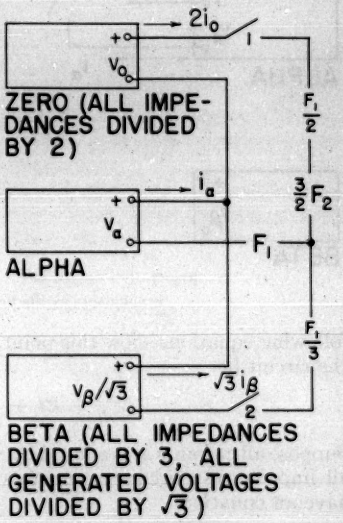


FIG. 10.

Equivalent circuits for fault representations.

considered, the causes of the fault may be lightning flashover, obstructions falling on the lines, or broken conductors; the cause of openings may be protector tubes, wind, natural extinction due to low fault currents or Petersen coil operation, or circuit breakers.

Line to ground fault and opening—phases a, n

See Fig. 7. Switch closes when fault occurs; opens when $i_a = 0$; $i_a = 0$; $i_0 = 0$.

Line to line fault and opening—phases b, c

See Fig. 8. Switch closes when fault occurs; opens when $i_b = 0$; $i_\beta = 0$.

Double line to ground fault changing to line to line fault and opening—phases b, c, n

See Fig. 9. Switches "1" and "2" close when fault occurs; switch "1" opens when fault changes to line to line; that is, when $i_b + i_c = 0$; $i_c = 0$; $i_a = 0$; switch "2" opens when $i_b = 0$; $i_\beta = 0$.

Double line to ground fault changing to line to ground fault and opening—phases a, b, n

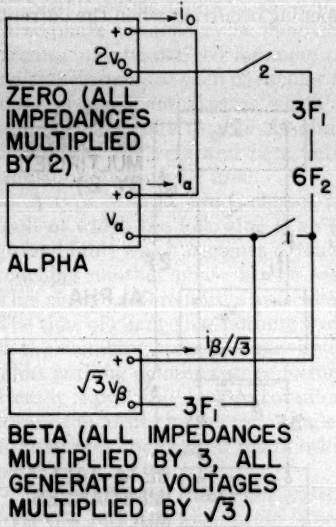


FIG. 11.

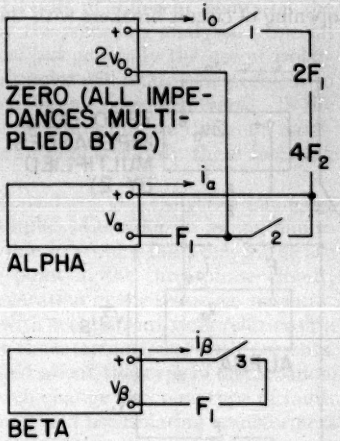


FIG. 12.

Equivalent circuits for fault representations.

See Fig. 10. Switches "1" and "2" close when fault occurs; switch "2" opens when phase b is cleared; that is, $i_b = 0$, $i_\beta = 0$; switch "1" opens when $i_a = 0$; $i_0 = 0$; $i_\alpha = 0$.

Double line to ground fault changing to line to ground fault and opening—phases b, c, n

See Fig. 11. Switches "1" and "2" close when fault occurs. Switch "1" opens when fault changes to line to ground fault on phase c; $i_c = 0$. None of the component currents pass through zero. However, the location of the switch is such that its current is $i_a + i_\beta/\sqrt{3}$ and this sum is zero when $i_c = 0$. Switch "2" opens when $i_b = 0$; $i_a = 0$; $i_\beta = 0$; $i_0 = 0$.

Triple line to ground fault changing to a double line to ground fault—phases a, b, c, n

See Fig. 12. Switches "2" and "3" close when fault occurs. When fault changes to a double line to ground fault on phases b, *c*, and *n*, switch "2" opens and switch "1" closes; $i_a = 0$; $i_a = 0$; $i_0 = 0$.

Conductor openings

Two cases will be considered here. Both will deal with sequential opening of circuit breakers with the opening occurring when the current

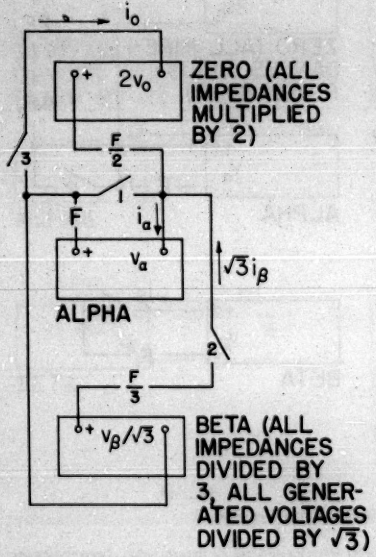


FIG. 13.

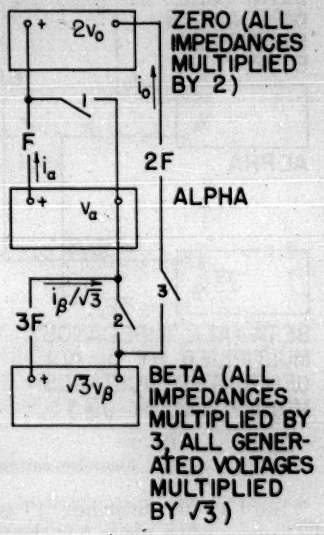


FIG. 14.

Equivalent circuits for fault representations.

passing through a pole of the breaker goes through zero. A direct opening of switches in the component networks is not possible since at the time of opening the currents in these networks will be finite. However, actually the current need not be interrupted; it is only necessary to shift it from one wire to another. A location of switches is possible which yields the correct equivalent circuits at the appropriate times and so that the current passing through them at the time of opening is zero.

See Fig. 13. Here the circuit is so arranged that openings in order phases *b*, *c*, and *a* can be handled. Originally all three switches are

closed. When $i_b = 0$, switch "1" opens; its current is $2i_0 + \sqrt{3}i_\beta - i_\alpha = 2i_b = 0$; switch "2" opens when $i_c = 0$; $i_\beta = 0$; switch "3" opens when $i_a = 0$; $i_0 = 0$; $i_\alpha = 0$.

See Fig. 14. Here the circuit is so arranged that openings in order phases a, b, and c can be handled. Originally all three switches are closed. When $i_a = 0$, switch "1" opens; its current is $i_0 + i_\alpha = i_a = 0$; when $i_b = 0$, switch "2" opens; its current is $i_\beta/\sqrt{3} - i_\alpha = 2i_b/3 = 0$; switch "3" opens when $i_c = 0$; $i_0 = 0$; $i_\alpha = 0$; $i_\beta = 0$.

Simultaneous faults or subsequent faults at different points on the three-phase circuit may be handled by the preceding analysis. Special circuits may be derived for some cases, but generally the use of isolating transformers at each particular fault point will reduce the problem to a set of single unbalances which have been considered here. If the equivalent circuits for the simultaneous faults require different modifications of the zero and beta networks then certainly these isolating transformers must be used.

A type of fault which cannot be considered generally is a changing fault of which the following is an example. Consider an arcing line to ground fault which becomes a double line to ground fault due to the arc touching another phase at the same point in the three-phase circuit. This change will require a switching operation in the sequence network. The time of change has nothing to do with zero current-time relationship so the switching operation may involve the interruption of finite currents. Thus nothing definite can be formulated about this type of disturbance. Here it is probably best to consider each change as a new type of fault, form equivalent circuits for a single fault, and use isolating transformers to maintain the desired relationships at the terminals of the sequence networks.

In the study of transient disturbances to three-phase circuits, another form of circuit representation is possible. This involves the setting up of the three-phase circuit in miniature on the calculator (33). The representation of the problem in Clarke components has the advantages of requiring less series line impedances and less line capacitances, using sequence quantities with which power systems engineers are more familiar, and making an analysis of transients and traveling waves a single-phase problem for which treatments appear in many textbooks.

CONCLUSIONS

1. An extension of the a-c. network calculator to the solution of transient problems occurring on three-phase power systems utilizing single-phase representation of the problem has been made.

2. Clarke components serve well in transient analysis and provide exact equivalent circuits for all types of transient disturbances. No troublesome switching operations or phase shifting transformers are required.

3. Representation by Clarke components requires less impedances than representation by a miniature three-phase system and may be simpler.

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(35) *Loc. cit.* 29, closing discussion by the authors.

APPENDIX. LONG LINE EQUATIONS IN CLARKE COMPONENTS

To derive the long line equations of three-phase lines in Clarke components, we follow well-established ideas (34). All circuit parameters are per unit length of line and the circuit is assumed to be symmetrical. For resistance and inductance the symbols of Eq. 13 shall be used. For capacitance and conductance let

$$\begin{aligned}
G_{an} &= \text{the conductance between phase and neutral wires,}\\
G_{ab} &= \text{the conductance between phase wires,}\\
C_{an} &= \text{the capacitance between phase and neutral wires,}\\
C_{ab} &= \text{the capacitance between phase wires,}\\
G_{11} &= G_{an} + 2G_{ab} = \text{the complete phase self conductance,}\\
G_{12} &= -G_{ab} = \text{the complete phase mutual conductance,}\\
C_{11} &= C_{an} + 2C_{ab} = \text{the complete phase self capacitance, and}\\
C_{12} &= -C_{ab} = \text{the complete phase mutual capacitance.}
\end{aligned} \tag{39}$$

Defining impedance operators we have

$$\begin{aligned}
z_{11} &= R_{11} + L_{11}\frac{\partial}{\partial t},\\
z_{12} &= R_{12} + L_{12}\frac{\partial}{\partial t},\\
y_{11} &= G_{11} + C_{11}\frac{\partial}{\partial t},\\
y_{12} &= G_{12} + C_{12}\frac{\partial}{\partial t}.
\end{aligned} \tag{40}$$

Then the canonical equations are

$$-\frac{\partial}{\partial x}\begin{bmatrix} v_a \\ v_b \\ v_c \end{bmatrix} = \begin{bmatrix} z_{11} & z_{12} & z_{12} \\ z_{12} & z_{11} & z_{12} \\ z_{12} & z_{12} & z_{11} \end{bmatrix}\begin{bmatrix} i_a \\ i_b \\ i_c \end{bmatrix} \tag{41}$$

and

$$-\frac{\partial}{\partial x}\begin{bmatrix} i_a \\ i_b \\ i_c \end{bmatrix} = \begin{bmatrix} y_{11} & y_{12} & y_{12} \\ y_{12} & y_{11} & y_{12} \\ y_{12} & y_{12} & y_{11} \end{bmatrix}\begin{bmatrix} v_a \\ v_b \\ v_c \end{bmatrix}. \tag{42}$$

Transforming to Clarke components

$$-\frac{\partial}{\partial x}\begin{bmatrix} v_0 \\ v_\alpha \\ v_\beta \end{bmatrix} = \begin{bmatrix} z_0 & 0 & 0 \\ 0 & z_1 & 0 \\ 0 & 0 & z_1 \end{bmatrix}\begin{bmatrix} i_0 \\ i_\alpha \\ i_\beta \end{bmatrix} \tag{43}$$

and

$$-\frac{\partial}{\partial x}\begin{bmatrix} i_0 \\ i_\alpha \\ i_\beta \end{bmatrix} = \begin{bmatrix} y_0 & 0 & 0 \\ 0 & y_1 & 0 \\ 0 & 0 & y_1 \end{bmatrix}\begin{bmatrix} v_0 \\ v_\alpha \\ v_\beta \end{bmatrix}, \tag{44}$$

where

$$\begin{aligned} z_0 &= R_0 + L_0\frac{\partial}{\partial t} \\ z_1 &= R_1 + L_1\frac{\partial}{\partial t} \\ y_0 &= G_0 + C_0\frac{\partial}{\partial t} \\ y_1 &= G_1 + C_1\frac{\partial}{\partial t} \end{aligned} \tag{45}$$

and

$$\begin{aligned} R_0 &= R_{11} + 2R_{12} \\ R_1 &= R_{11} - R_{12} \\ L_0 &= L_{11} + 2L_{12} \\ L_1 &= L_{11} - L_{12} \\ G_0 &= G_{11} + 2G_{12} \\ G_1 &= G_{11} - G_{12} \\ C_0 &= C_{11} + 2C_{12} \\ C_1 &= C_{11} - C_{12}. \end{aligned} \tag{46}$$

Ionization Control Makes Indoor Air Like Outdoor Air.—Sufferers of asthma, chronic colds and hay fever may, at last, find some relief by using new ion controllers which make indoor air like outdoor air at its best. This latest aid was introduced by the Wesix Electric Heater Company of San Francisco and will be marketed on a commercial scale in the near future.

Although the medical profession has known for some time that negatively ionized air breathed by humans has a therapeutic effect on certain ailments, it has only been recently that scientists have developed a method of providing negative ions without objectionable ozone and nitrous oxide.

Devices known as Ionix and Wexon units are ion emitters arranged to filter out positive ions while negative ions are carried into the room by air currents passing through the units. The principle of the heater is simply that stale air is drawn in through the bottom of the unit and the ion rectified air is released through the top.

Research studies were commenced on this project in 1939 at Stanford University Electrical Engineering School of Palo Alto, in conjunction with engineers of Wesix Electric Heater Company. Tests have shown that the new type of Wesix radiant convection type heaters, operating on ordinary house current, generate these necessary ions without harmful effects. A simple means for screening out positive ions and selecting negative ions for distribution in the household has also been developed by Wesix engineers. One of the principal components of the heater is a pure nickel cloth which forms a negatively charged grid surrounding the heating element which is made of nickel-chromium alloy.

Very stimulating results have been uncovered by the Wesix Electric Heater Company's research program in this field. The project has uncovered so many unlooked-for effects that the company is now sponsoring controlled clinical investigations at five recognized universities and medical schools throughout the country for the purpose of determining the scientific basis from the results obtained.

These Wesix ionizers are adaptable to all types of climate. For cold weather conditions the Ionix heater is rated 1000 Watts or more. In warm climates the Wexon is a 30-Watt platinum high-temperature filament or equivalent emitter, arranged to generate negative ions when heat is not needed. In this type of heater, the negative ions neutralize positively ionized dust and pollen as air is circulated through the unit. The Wesix Electric Heater Company has perfected this simple means to make this formerly overlooked negative ions therapy available to everyone.

A revolutionary free-piston power plant which combines conventional diesel engine thermal efficiency with the advantages of gas turbine operation is described in a report available from Office of Technical Services, Dept. of Commerce. Supplementing previously published information on this French developed engine, the report deals with the problem of obtaining operating stability, achieved despite the absence of con-rod and crankshaft linkages. The free-piston gas producer is actually a supercharged 2-cycle diesel producing exhaust gas at 75 psi. pressure and 900-950° F. which generates power by expansion to atmospheric pressure in a gas turbine. Order PB 100 124 "Operating Stability of Floating Piston Gas Generators," $1.75 microfilm, $2.50 photostat, from Library of Congress Publication Board Project.

NOTES FROM THE NATIONAL BUREAU OF STANDARDS*

NOISE FIGURE STANDARDS

The noise figure, a fundamental measure of the quality of linear electrical networks, is of basic importance in radar, telemetering, and all communications. In these systems some of the limitations on reliability, sensitivity, and distance are set by the type and magnitude of noise in the device as well as by the noise produced ahead of the network input terminals. In order to assist laboratories and industry in the evaluation of this important factor, the National Bureau of Standards is offering a calibration service for the noise figure in the frequency range of 500 Kc. to 30 Mc. Standards for this purpose have been developed by M. Solow, I. W. Hammer, and P. H. Hass of the Bureau's Central Radio Propagation Laboratory.

The noise figure of a linear network is the ratio of the available noise power at the output (the total network and source noise) to the available noise power at the output source alone. Noise power in an electrical network (that is, a radio receiver) is generated by the network resistance (Johnson noise) and its vacuum tubes (shot noise). The noise figure is a function of frequency and of the network and source impedance; both are measurable to a high degree of accuracy by various independent techniques. With these important parameters of a unit accurately evaluated, calibration can now be made. Thus, the Bureau's calibration method breaks down into one involving only five components—a temperature limited noise diode, a two-terminal source network, a four terminal network under calibration, an attenuator, and a sensitive voltmeter.

The equivalent noise resistance used in evaluating the technique utilizes the concept that the noise power from the network can be represented by the Johnson noise of this resistance. Experimental verification of this theory was made with eleven different values of test impedance and at frequencies of 0.5, 4.3, 12, and 30 Mc. Measurements made with the temperature-limited diode conclusively proved that the equivalent noise resistance was constant for all the values of test network impedance at each frequency. The evaluation has shown that the Bureau's calibration method will yield precise noise figures. It was also proved that this method of calibrating noise figures is valid for a matched or unmatched condition of input impedance. In addition this method may be successfully applied to measuring the impedance of two-terminal networks.

* Communicated by the Director.

Thus, with the theory verified by a wide range of experiments, the National Bureau of Standards is equipped to provide standards for noise figure measurements. Calibrations can be made for high gain, linear, four-terminal networks such as radio recievers and amplifiers (10 to 150 ohms) source impedances to ± 0.2 decibel and at frequencies up to 30 Mc. Work is still in progress to extend noise figure standards to 300 Mc.

NEW LABORATORY FOR DURABILITY STUDIES OF PROTECTIVE COATINGS

A special laboratory, recently designed and constructed by the Building Technology Division of the National Bureau of Standards, eliminates the major difficulties that have in the past led to anomalous results in the accelerated durability testing of exterior covering materials. Developed primarily for studying organic coating materials such as asphalts, tars, and paints, the new facilities are sufficiently flexible to meet the requirements of most materials exposed to weathering.

Materials intended for outdoor use are frequently exposed in the laboratory to regular cycles of artificial sunlight, heat, and water for the purpose of hastening their deterioration. Although such exposure—called accelerated weathering—is generally accepted as a means for determining the relative resistance of similar materials to weathering, results obtained by different laboratories, and even by the same laboratory, do not always show good agreement. This lack of correlation results from variables inherent in the testing procedures that have in the past been largely uncontrolled. These include solids content of the water supply, fluctuations in the light intensity, and irregular operation of the system due to failure of one or more functions while the others continue to operate.

Automatic control of all functions, and of the temperature and purity of the water within narrow limits, assures uniform operation at all times of the Bureau's improved accelerated weathering laboratory. The room 16 by 25 ft. in size, is maintained within ± 1° C. of any desired working temperature, at present, 27° C. The relative humidity is not controlled directly, but remains in the range of 40 to 60 per cent as a result of the subcooling of fresh air brought into the room. The floor area is divided into three portions: a central working space and two areas enclosed in hoods to house the testing machines.

In one hood are six machines built at the Bureau to utilize the enclosed carbon arc as a source of light. The operating range of each light is regulated at 130 to 145 volts and 15 to 17 amperes by means of chromel C ballast resistors in series with the arc. An aluminum drum 31 in. in diameter and holding sixty exposure panels (2¾ by 6 in.) is rotated around each arc at a rate of one revolution per minute. Air at ambient conditions is circulated around the panels at a low velocity until they reach a predetermined maximum temperature. At this time a thermo-sensitive

element mounted in one of the panels turns on the high speed blower, which materially increases the air flow and thereby reduces the panel temperatures. Deionized water, controlled by a cycle clock, can be sprayed on the test panels for any portion of their exposure period.

The heart of each machine is a time-delay relay which turns off all functions of that machine if the arc should cease to operate for a period in excess of 30 seconds. Voltage and current are continuously indicated while the running time and power consumed by the arc are recorded by suitable instruments on each machine. Wiring circuits have been designed to take advantage of the 2 per cent voltage control supplied the Bureau.

The other hood encloses a modified commercial open-drum Atlas Weatherometer, two twin-arc Weatherometers, and an Everready X-1A machine. These are controlled in a manner similar to that of the other machines.

An air conditioner of tremendous size could be required to control the temperature of the space around all 10 machines, which dissipate 39 KVA of energy. This temperature is therefore regulated by blending various amounts of outside corridor air (ranging from 15° to 35° C.) with conditioned air at 25° ± 1° C. The resultant mixture is distributed to the machines in each hood by a 2000 cfm. blower with a proportioning damper in its intake plenum. The blended air is distributed in turn to the vicinity of each machine by aluminum ductwork. Further conditioned air is drawn through a two-inch opening running along the floor of each of the hoods by an exhaust blower on the hoods which removes about 30 per cent more air than is delivered by the ducts. In this way, the ambient air at each machine is maintained at 29 ± 4° C.

Water-conditioning equipment, located on a balcony above one of the hoods, delivers recirculated, deionized water with about 1 ppm. of dissolved solids at any desired temperature between 4° and 25° C. (± 1° C.) to the six machines in that hood. At present, to introduce a cyclic thermal shock to asphalt exposure panels, the equipment is operating at 5° C. and 25 psi. The pH of the water is maintained by the deionizer at 6.2. Water for the other machines is delivered at room temperature and a pH of 6.0 by another deionizer. All machines can be operated with tap water if deionized water is not required.

Because accelerated durability testing of exterior coverings does not at the present time permit a quantitative correlation with actual weathering, the National Bureau of Standards has constructed a set of exposure decks on the roof of the laboratory that will hold 3000 test panels. Materials are thus exposed concurrently on the roof and in the controlled laboratory. Data from these studies should provide some statistical correlation between outdoor weathering in the Washington, D. C., area and in accelerated durability testing in each of the several types of machines.

STANDARDIZATION OF IRON-CONSTANTAN THERMOCOUPLE TABLES

The iron-constantan thermocouple is widely used in industry for measurement of temperatures up to 1500° F. Yet it is the only type of thermocouple for which uniformity of commercial practice has not been established in this country. Recently, at the request of the Scientific Apparatus Makers of America, the National Bureau of Standards has undertaken to derive a temperature-emf. reference table for iron-constantan thermocouples which would be suitable for adoption as a standard.

The temperature-emf. tables for thermocouples which the Bureau has issued from time to time have been used extensively in science and industry, not only to convert thermocouple voltages into the equivalent measured temperatures but also in the preparation of purchase specifications for thermocouple wire and in defining the relation between impressed emf. and scale reading for pyrometers which indicate temperature directly. In practice the calibration of an individual thermocouple is ordinarily given in terms of deviations from such a reference table. Since it is usually not convenient to change the scale of a direct-reading pyrometer to compensate for variations in the thermocouples which may be used with it, thermocouples are commonly purchased under specifications which ensure that they will correspond closely to the graduation of the particular model of pyrometer used.

Some time ago, an investigation[1] of iron-constantan thermocouples by the Bureau disclosed that the thermocouples of this type which were then commercially available did not follow closely any of the several temperature-emf. relations for iron-constantan in use by pyrometer manufacturers. The Bureau therefore derived a new table, usually known as the RP 1080 table, which represented the average behavior of the iron-constantan thermocouples on the market. However, rather than alter the scales of their pyrometers to conform to the actual properties of the thermocouples then being sold, most of the various manufacturers made efforts to develop and make available thermocouples which would correspond to the temperature-emf. tables they were using. As a result, a number of widely diverse reference tables for iron-constantan thermocouples are still in use.

In an effort to achieve greater uniformity, the Recorder-Controller Section of the Scientific Apparatus Makers of America, representing several of the pyrometer manufacturers, in 1948 adopted the temperature-emf. relation known as the 1913 (L and N) table[2] as a tentative standard for iron-constantan thermocouples over the range from 32° to 1400° F. This relation was chosen because it is in wider use than any

[1] WILLIAM F. ROESER AND ANDREW I. DAHL, "Reference Tables for Iron-Constantan and Copper-Constantan Thermocouples," *J. Research NBS*, Vol. 20, p. 337 (1938), RP 1080.

[2] Identical with the tables given in "International Critical Tables," Vol. 1, p. 59 (1926) and in *Tech. Pap. BS*, Vol. 14, p. 306 (1920–21), T170, Table IV, Col. L.

other. However, experience has shown that it is apparently not possible to produce a thermocouple which will correspond precisely to the 1913 table over the entire range from 32° to 1400° F. Moreover, any thermocouple which closely matches this table in the range below 1400° F. will deviate greatly from it in the range above 1400°F. The SAMA therefore requested the National Bureau of Standards to derive a reference table which would be free from these defects. It was felt that such a table should be as close to the 1913 table as possible, yet should represent the properties of an actual commercially available thermocouple.

Eight different lots of matched iron and constantan wires were submitted to the Bureau by four members of SAMA. The thermal emf.'s of the various wires *versus* platinum were measured in the Bureau's pyrometry laboratory from 32° to 1800° F. From these data, the corresponding thermal emf.'s for the iron-constantan thermocouples were computed. In addition, chemical and spectrographic analyses of the iron samples and hardness determinations were made. It was found that the spread in the thermal emf.'s of the iron-constantan thermocouples below 1400° F. was not large compared to usual commercial tolerances, even though the iron samples differed among themselves by as much as 0.8 millivolt at 1400° F. (equivalent to about 20 degrees in the indication of an iron-constantan thermocouple) and the constantan samples differed similarly. Present preference is to base the new SAMA table upon the temperature-emf. relation of thermocouple A-1, since this thermocouple deviates least from the 1913 table in the range of normal use.

The question may now be raised whether the field of use of this proposed new table can be broadened through industry-wide acceptance. Support by this Bureau for any new table (taking the form of official publication and possibly resulting in inclusion in a Federal Specification) might conceivably cause hardship to suppliers and users of iron-constantan thermocouples who have based their operations on the RP 1080 table. Thus, it is felt that the views of all interested parties should be obtained as a prerequisite to the adoption of this table. The questions that naturally occur if the proposed table can be adopted are the following: Is industry-wide standardization on a single reference table for iron-constantan desirable? Is the 1913 table, or a close approximation to it, acceptable for this purpose? What degree of hardship will be occasioned by abandonment of the RP 1080 table? The reaction of producers and users of pyrometers is a matter of interest to the Bureau, and replies to the above questions as well as related comments can be addressed to the National Bureau of Standards, Temperature Measurements Section, Washington 25, D. C.

THE FRANKLIN INSTITUTE

MINUTES OF STATED MEETING

April 18, 1951

The April Stated Meeting of The Franklin Institute was held in the Lecture Hall on Wednesday evening, April 18, 1951.

The President called the meeting to order at 8:15 P.M. He reported that the minutes of the Stated Monthly Meeting for March were printed in full in the April issue of the JOURNAL, and there being no corrections or additions, they were approved as printed.

The Secretary reported the following elections to membership for the month of March:

Active	36
Associate	29
Student	9

The total Institute membership as of 31 March was 6,027.

The speakers of the evening, Commander and Mrs. Finn Ronne, were introduced by the President. Commander Ronne is a member of U. S. Naval Reserve and author of "Antarctic Conquest." A briiliant all-color film lecture of their famous fifteen-month polar exploration of 1947–48 was presented by the speakers. The colored film illustrated the beautiful scenery and wild life of that region. From the scientific standpoint, Commander and Mrs. Ronne spoke enthusiastically about their scientific observations, and in particular, those made for Dr. W. F. G. Swann of the Bartol Research Foundation on Cosmic Radiation.

Commander and Mrs. Ronne were given a rising vote of thanks as the meeting adjourned.

HENRY B. ALLEN
Secretary

JOURNAL OF THE FRANKLIN INSTITUTE

The following papers will appear in the JOURNAL within the next few months:

DRUCKER, D. C. AND F. EDELMAN: Some Extensions of Elementary Plasticity Theory.

EPSTEIN, HERMAN: Solution of Transients in Active Four-Terminal Networks.

MOON, PARRY AND DOMINA EBERLE SPENCER: Interflectance Calculations for Various Luminaires.

SILBERSTEIN, LUDWIK: The Effect of Gradual Light Absorption in a Film of Non-Negligible Thickness.

WEINBERG, LOUIS: Solutions of Some Partial Differential Equations (with Tables).

HUNICKE, A. BYRON: The Valved-Tremie Applied to Subaqueous Concrete Structures.

RAMSEY, W. E.: The Slow Discharge in a Non-Self-Quenching Geiger Counter.

HIGGINS, THOMAS J.: Solution of Boundary Value Problems by Multiple Laplace Transformations.

COMMITTEE ON SCIENCE AND THE ARTS

(Abstract of Proceedings of Stated Meeting held Wednesday, April 11, 1951.)

HALL OF THE COMMITTEE,
PHILADELPHIA, APRIL 11, 1951.

DR. RUPEN EKSERGIAN in the Chair.

The following reports were presented for final action:

No. 3195: Collins Helium Cryostat.

This report recommended the award of a John Price Wetherill Medal to Samuel C. Collins, the inventor, and of an Edward Longstreth Medal to Howard O. McMahon, his coworker in subsequent development, both of Cambridge, Massachusetts, "In consideration of the invention and subsequent development of the first liquefier of helium to operate without the aid of external refrigerants, an admirably designed, reliable machine which is produced in quantity and has materially increased the number of laboratories and persons engaged in work at ultra low temperatures."

No. 3230: Hydro and Oleopneumatic Accumulator.

This report recommended the award of a Certificate of Merit to Jean Mercier, of Paris, France, "In consideration of new and important combinations of principles already known and for many inventions covering the details of the oleopneumatic accumulators."

No. 3234: Clamer Medal.

This report recommended the award of the Francis J. Clamer Medal to John Chipman, of Cambridge, Massachusetts, "In recognition of his contributions as an individual and as a teacher to the application of the theories of physical chemistry to steelmaking practice."

JOHN FRAZER,
Secretary to Committee.

MEMBERSHIP

ACTIVE MEMBERS ELECTED AT THE MEETING OF THE BOARD OF MANAGERS

April 18, 1951

ACTIVE FAMILY

H. Richard Dietrich
Byron H. Edwards
Alfred Green
Connie Lenz
John G. Pew, Jr.
Richard J. Street

ACTIVE

William M. Adams, Sr.
Nolan N. Atkinson, M.D.
J. Bruce Bredin
Clarence C. Briscoe, M.D.
Geoffrey R. Buckwalter
John B. Derr
John U. Druck
Laurence A. Fischer
Charles H. Fritz
Andrew C. Fulton
Samuel A. Gould
Mozam Hosein, M.D.
Kitty Lus
A. Herbert Marbach, M.D.
Hugh G. Oldach
Sydney C. Orlofsky
David Osterneck
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F. Leon Yetter

ACTIVE NON-RESIDENT

Thomas Maier

NECROLOGY

George A. Pierce '13

MUSEUM

The issue of the JOURNAL for May, 1851 (exactly one hundred years ago) contains a description by Leon Bernard Foucault of the first demonstration of the diurnal turning of the earth.

First proposed in reasonable form by Aristarchus about 230 B.C., the heliocentric theory with a rotating earth was put on a logical basis by Copernicus in 1543. By the time of Galileo's death, 99 years later, a bare majority of educated men believed it. Actual astronomical proofs of the motion of our planet around the sun were found in 1729 (aberration) and in 1838 (annual parallax).

Direct experimental proof of the daily rotation of the earth has, however, always been disconcertingly, almost embarrassingly, hard to secure. Any sane effort to explain the alternation of night and day, coupled with other simple astronomical phenomena demands such a motion, and conviction necessarily follows acceptance of the law of universal gravitation. Virtually every advance in astronomy piles up more evidence that Copernicus was right.

To exclude observation of celestial bodies and still demonstrate the rotation is quite another question, one which stumped the experts for centuries only to find a practical solution just one hundred years ago.

> "On the 3rd of February, 1851, M. Foucault communicated to the French Academy (*Comptes Rendus*, **32**, 135) an experiment which has attracted more public attention than perhaps any other of modern times, namely his celebrated experiment of the pendulum. The phenomena of astronomy have long since proved the rotation of the earth; but we, living on it, are not sensible of that rotation. In one experiment had the effect of the rotation been manifested, namely, in the eastern deflection of a body let fall from a great height. But in this case the experiment could only be attempted in particular places (mines) and the result was only a small deflection which was liable to be interfered with by a variety of disturbing causes. But M. Foucault showed that a pendulum suspended so as to oscillate in all vertical planes alike, withdrawn from the vertical and then left to itself, became by the very fact of its motion in some sense a celestial body; and while the plane of the motion tended to remain parallel to itself, and would actually do so at either pole, the earth turned round under it, and thus the plane of motion was affected with an apparent rotation in the direction, in our hemisphere, of the hands of a watch, the velocity of rotation decreasing from one turn in 24 hours at the pole to zero at the equator, in proportion to the sine of the latitude and changing sign in passing into the southern hemisphere." (Lord Wrottlesley, President of the Royal Society in presenting the Copley Medal to Foucault in 1855.)

The experiment by which Foucault had demonstrated visually the earth's daily rotation was demonstrated in the Pantheon in Paris, and has since been repeated by a number of others. It is performed daily in The Franklin Institute Museum where an 1800-lb. ball swings on the end of an 85-ft. bronze wire, one of the most precisely hung large Foucault pendulums on display in the world. Each day it is started swinging in a north-south direction. In a very few minutes it can be seen to be working off to the east of north and the west of south, convincing proof that the earth is turning under it. Or is it quite as convincing as it should be?

In the first place the pendulum does not make a revolution in 24 hours though it is heavy enough and free from friction to keep swinging for 30 hours or more. It only goes through 230 degrees, because it is in Philadelphia (Latitude 40 degrees north).

To "make it come out right" the entire pendulum and support would have to be installed in an absolutely rigid tower at the South Pole (the North Pole would not do—nothing but drifting ice). A perfect pendulum would here appear to rotate the full 15 degrees per hour, in fact it would complete a full revolution in 23 hours, 56 minutes, and 4 seconds; that is, in one sidereal day. This shows that the natural frame of reference in our immediate part of the universe is the positions of the fixed stars.

If it is inconvenient to go all the way to the South Pole, let us look at the other mechanical proofs of the earth's rotation. First, there is the slight deviation to the right (in the Northern hemisphere) of projectiles. Second, is the eastward deviation of freely falling bodies. Third, is the polar axis-seeking behavior of a gyro-compass. In the first two cases the deviation is so slight as to be disentangled only with difficulty from unavoidable causes of far larger "scatter." The last requires an understanding of the theory of gyroscopic precession. There is still a more elegant proof best presented in 1915 by Arthur H. Compton, but this involves controlled conditions and observations with microscopes.

It is clear that for public demonstration the Foucault pendulum gives the best experimental means that man has yet devised to show that the earth really does spin. If conviction does not come with contact and contemplation the moral may be that we should not look inward to see the world as it really is. Our planet is but a member of the vast solar family and it is to the sky that we must look to see the part it plays in a greater harmony.

LIBRARY

The Committee on Library desires to add to the collections any technical works that members would wish to contribute. Contributions will be gratefully acknowledged and placed in the library. Duplicates received will be transferred to other libraries as gifts of the donor.

Photostat Service. Photostat prints of any material in the collections can be supplied on request.

The Library and reading room are open on Mondays, Tuesdays, Fridays and Saturdays from 9 A.M. until 5 P.M.; Wednesdays and Thursdays from 2 P.M. until 10 P.M.

On and after June 1, the Library hours will be 9 A.M. until 5 P.M. on Mondays, Tuesdays, Wednesdays and Fridays; 2 P.M. until 10 P.M. on Thursdays; and 9 A.M. until 12 noon on Saturdays.

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THE FRANKLIN INSTITUTE LABORATORIES FOR RESEARCH AND DEVELOPMENT

NATIONAL ASPHALT RESEARCH CENTER

The National Asphalt Research Center[1] was formally established by The Franklin Institute on January 29, 1951, as a part of its Laboratories for Research and Development. The opening symposium was attended by thirty-three representatives of the current twenty-three Sponsors engaged in the production and use of asphalts and allied materials such as oil, roofing, molded products, linoleum, and sealing compounds.

[1] Hereafter referred to as NARC.

The idea of establishing a national research center at The Franklin Institute arose from a friendly discussion among asphalt technologists concerning the needs of the industry. The Institute was very receptive to the idea, having recognized the need for fundamental asphalt research from the problems it had faced in developing better asphalt products for industrial sponsors, and the enthusiasm engendered by a paper[1] on the basic properties of asphalts presented by a staff member at the Detroit meeting of the Association of Asphalt Paving Technologists in February, 1949.

The organization of the NARC is simple and effective. The sponsoring companies guide the work through a steering committee on which each company has one voting representative. The work is controlled and performed by The Franklin Institute Laboratories for Research and Development, and the methods and results are published for the Sponsors in the form of confidential reports. Each phase is broken down into tasks which can be completed and reported at relatively short intervals, and a brief progress report is issued each month.

FIG. 1. Tester for measuring rheological properties of asphalts.

The findings are the property of the Sponsors. The flow of ideas concerning the work of the NARC among the personnel of the Institute and the Sponsors will be facilitated by correspondence and visits between steering committee meetings. Furthermore, the extensive bibliography which the Institute is preparing on asphalt and related subjects will be available to Sponsors.

The staff responsible for the work of the NARC is not only familiar with asphalt and asphalt products, but also with principles and techniques in a wide range of physical sciences, mathematics, and engineering. A modern laboratory has been equipped for and adapted to the study and testing of asphalt and asphaltic materials. The entire facilities of the Laboratories are available for use as needed. Some new equipment was procured and the tester for measuring rheological properties of asphalts (Fig. 1) is being developed by the Laboratories. It is but one of many specially designed pieces of equipment which will be used in attaining the objectives of the NARC.

[1] "Physical Testing of Asphalts and Similar Materials," by Edmund Thelen. Mr. Thelen is Project Leader of NARC and Chief, Colloids Section, The Franklin Institute Laboratories for Research and Development, Philadelphia, Pa.

Basic conceptions of the NARC were perfected and crystallized in discussions with many asphalt technologists and industrial leaders throughout the country. They are resolved as follows:

a. The asphalt characteristics which are of importance to virtually all makers and users of asphaltic products depend upon relatively few basic properties. Basic work on these characteristics at a non-competitive level can be carried out most effectively through cooperative effort. These characteristics are described in such terms as:

Viscosity	Brittleness	Wetting of Filler
Sag	Toughness	Age-hardening
Slide	Impact Resistance	Stability
Float Viscosity	Cold Flow or Creep	Compatibility
Penetration	Vibration Resistance or Fatigue	Syneresis
Ductility	Adhesion of Granules	Gelation in Cutbacks
Stickiness	Adhesion to Stone	Stain
Tack	Joint Strength	Stripping
Pliability	Water Penetration	Weathering
Tensile and Other Strengths	Abrasion Resistance	Film Properties
Pouring Properties		Temperature Susceptibility
		Sensitivity to Light

b. These characteristics are all functions of a relatively few fundamental properties at various temperatures, rates of shear, and environments. Hence, a field of the NARC is (*a*) to define these basic properties, both chemical and physical, in scientific terms; (*b*) to develop practical and efficient ways for measuring them; and (*c*) to correlate them with the various engineering characteristics which they control. As this work is performed it will be possible, from comparatively few basic tests, to determine the suitability of an asphalt not only for present uses, but for new uses.

c. The fundamental properties of an asphalt depend upon its chemical composition and colloidal structure, which in turn depend upon the raw materials used and their methods of manufacture. While it will be possible to make many deductions regarding colloidal structure and chemical composition from measurements of fundamental properties, a recognized aim of the NARC is to develop ways for describing and measuring the important composition variables.

In essence, the work of the NARC is to construct rigorous logical bridges to connect engineering requirements of applications to fundamental properties and fundamental properties to colloidal structure and composition. A scientific technology for asphalts is complete when the formulators take the comparatively easy step of correlating the raw materials and processing to the colloidal structure and composition. In practical terms, the purpose of the NARC is to develop conceptions and methods that will be used by the sponsoring companies as tools for creating better asphalts and asphaltic products, and for extending the use of asphalts in new applications.

BOOK REVIEWS

JOSIAH WILLARD GIBBS: THE HISTORY OF A GREAT MIND, by Lynde Phelps Wheeler. 264 pages, port., 16 × 24 cm. New Haven, Yale University Press, 1951. Price, $4.00.

After his student days were ended Josiah Willard Gibbs' interests were directed to railroad car brakes and the governor of the steam engine, but this engineering phase did not last long. It became submerged under the theoretical treatment of his subject when once the science of thermodynamics opened before his penetrating mind. Without discussing his thoughts with other scientists, Gibbs meditated long before issuing his first paper, "On the Equilibrium of Heterogeneous Substances," in which he developed the celebrated "phase rule." The significance of this paper eluded contemporary American scientists, who could see no genius in its abstractions. Fortunately, European reaction was different. Clerk Maxwell and Ostwald saw within the paper the fundamentals of a brilliant new attack upon several perplexing problems.

Then followed a period in which Gibbs developed vector analysis and multiple algebra before he placed the cap-stone on his career with the publication of "Elementary Principles in Statistical Mechanics." It was fortunate for Gibbs that he attached no importance to the question whether the matter he chose to discuss had any significance for other people. Otherwise he might have been discouraged, for his work failed to win any immediate appreciation. A few men, like Planck, Ornstein, and Lorentz, perceived its value and, while not in full agreement with all the author's views, they were willing to advance along the line of thought he indicated. Not until the advent of wave mechanics and the quantum theory did it become apparent that Gibbs had made valuable contributions to the realization of entropy.

There are several reasons why Gibbs' work made only slight impressions upon contemporary American thought. Dr. Wheeler deals skillfully with these and, having shown that the neglect of Gibbs was largely based upon incorrect assumptions, points to the true reason that Gibbs was far too far in advance of his times. Recognition of his genius had to await the growth of physical chemistry as a profession.

However, Gibbs' work presents difficulties (he only offers the barest mathematical outlines of treatment) even to the science student with modern training, and a good interpretation is needed. Dr. Wheeler, a student under Gibbs and a former Associate Professor of Physics at Yale, has essayed the task and accomplished it with marked distinction. Some portions of the book will present difficulties to readers who lack a training in physics, but the subject does not lend itself to "popular" treatment. A comprehension of the greatness of this giant of the classical age of physical science cannot be gained without a knowledge of the field in which he worked.

Dr. Wheeler treats of Gibbs' work as a scientist and a teacher with sympathetic understanding and, without entering upon any exhaustive biographical detail, leaves a lasting impression of a simple kindly man, far removed from the picture of the remote dweller in an ivory tower. Not the least interesting portions of the book are those which deal with correspondence Gibbs conducted with the relatively few men who understood his work. This corrects the false impression that he was without any appreciation in his time. The author shows, too, that power and fertility of Gibbs' achievement have grown with time. It is a wholly satisfying study of a scientist of the first rank, whose life and work deserve the wider public this book should bring him.

T. COULSON

THE BEHAVIOR OF ENGINEERING METALS, by H. W. Gillett. 395 pages, illustrations, 15 × 24 cm. New York, John Wiley & Sons, Inc., 1951. Price, $6.50.

A very commendable purpose forms the basis for "The Behavior of Engineering Metals": To present, with a minimum of technical metallurgical terminology, the fundamental behavior of metals and how metallurgical practice controls their chemical and physical properties. Dr. Gillett, late Chief Technical Adviser of the Battelle Memorial Institute, has accomplished this in a very interesting and highly recommended book. For engineers and designers who are concerned with the behavior of metals, whether it be the yield strength, the elastic limit,

the fatigue strength, hardenability, machinability or ease of fabrication, the book admirably bridges the gap between metallurgy and practical engineering. It gives, in simple generally non-technical language, not only the meaning of many metallurgical practices and terms of which the designer has heard but seldom really knows, but also a sort of over-all picture of the whole field of design: the service conditions, expendability, life expectancy, seriousness of failure, specifications and test methods, alternate materials, cost and conservation. It could never be called a dull book.

Introducing his subject with a brief but clear description of such common metallurgical terms as cooling rate, sluggishness, austenite, martensite, tempering, hardenability and physical properties, and a discussion of the conventional mechanical tests for physical properties, Dr. Gillett proceeds to general design considerations, specifications, and limitations of conventional test methods—a very worthwhile and informative section. Chapters on stress concentration, on the statistical approach in the acceptance of material, and on the chemical and physical properties of steels follow.

These initial chapters are rather general. The author, in succeeding chapters, discusses the important commercial metals and alloys, with consideration of their selection for various uses. Wrought steels, heat-treated steels, cold-worked metals, ferrous castings, castings in general, light wrought alloys, common and not so common heavy metals, are each treated in some detail. Their production, composition, use and peculiar properties are frankly delineated. The latter part of the book concerns such special subjects as the effect of hydrogen inclusion on metal behavior, machinability, clad and coated metals, metal films and surfaces, powder metallurgy, bearing metals, wear and corrosion, and service at high temperatures.

Frequent quotations of general and test data from reliable sources enhance the authenticity of the book, and the numerous references following each chapter make it, in itself, a valuable source book. While the volume may not be particularly technical—and its avowed purpose is just that—it certainly contains a wealth of practical metallurgical, engineering and general design information.

E. W. HAMMER, JR.

FOUNDATIONS OF AERODYNAMICS, by A. M. Kuethe and J. D. Schetzer. 374 pages, illustrations, diagrams and tables, 15 × 23 cm. New York, John Wiley & Sons, Inc., 1950. Price, $5.75.

The authors state that the aim of this book is "To present the foundations on which aerodynamics rests in a form suitable for classroom use . . . and to give an understanding of the basic concepts rather than a complete coverage of methods and data." Since the field of aerodynamics has expanded so rapidly in recent years, there has been a crying need for a book of this sort and as such, this present work is welcomed, especially by those without a prior knowledge of the subject at the senior and graduate levels.

This book will not fully satisfy the specialists in mathematics, compressible fluids, fundamental theory, practical design, etc., since certain important data and subjects, of necessity, have been eliminated. This is justified, however, since in order to keep abreast of all phases of aerodynamics, one must have an up-to-date perspective of the entire field, and to give all the subjects the space they would ordinarily require by themselves would be inconsistent with the objective of the book which is further implied by its title, "Foundations of Aerodynamics."

The text offers a comprehensive background for the solution of the many practical problems in that fundamental details are treated completely enough by themselves and it is unique in several noteworthy respects. It is divided into three parts: subsonic fluid dynamics, compressible flow, and viscous flow. Careful consideration is given to an understanding of theory, its validation and limits of application. Eighteen chapters are included, the first of which treats the fluid medium and the formulation of the fluid problems. The next five chapters deal with the kinematics and dynamics of a fluid field and their application to the prediction of the characteristics of thin airfoils and the finite wing.

Chapters 7 through 11 concern the introduction to compressible inviscid fluids and the application to channels and airfoils in compressible flow.

Chapters 12 through 16 contain an introduction to the dynamics of viscous fluids, laminar flow in tubes and boundary layers, turbulence, turbulent flow in tubes and boundary layers, and the transition from laminar to turbulent flow.

Chapter 17 deals with certain aspects of viscous compressible flow and discusses the effects of viscosity on the flow of a compressible fluid and the alterations to the boundary layer due to compressibility. The last chapter presents a relatively short but concise application of the treatment of the previous chapters to the design and characteristics of wings at low subsonic, high subsonic, transonic and supersonic speeds.

As an added "plum," rightly divorced from the rest of the book, two appendices supplement the text. These appendices deal with a treatment of dimensional analysis and a development of the equations of motion and energy for unsteady compressible viscous flow, usually known as the Navier-Stokes equations.

S. M. BERKOWITZ

MECHANICS OF MATERIALS, by Harry D. Conway. 325 pages, illustrations, 15 × 24 cm. New York, Prentice-Hall, Inc., 1950. Price, $4.75.

Another book has been added to the long list of introductory volumes on the strength of materials. Differing little from the usual treatment of the subject, it does, however, possess several noteworthy features. Professor Conway of Cornell University, realizing a frequent lack of knowledge in later professional practice, regarding the validity of strength of material formulas applied to various structures, has particularly emphasized the assumptions upon which the formulas are based and their limitations. By including many illustrative problems, he has also sought to facilitate the transition from the theory of the classroom to the practice of the professional world. Included, also, is the stress analysis of thick cylindrical and spherical shells, not often found in elementary texts on strength of materials.

Marred occasionally by a clumsy style, the book is yet refreshingly concise and definite, and embraces a rather thorough discussion of the mechanics of materials. Commencing with the usual introduction into stress and strain, the author presents the general topics of torsion, shearing, bending and deflection of simple and indeterminate beams, stresses in beams, columns, statical indeterminacy and Castiglioni's theorem and failure under combined stresses. Welded and riveted joints are discussed in a separate chapter. An appendix contains moments of inertia of various sections.

E. W. HAMMER, JR.

HARVARD CASE HISTORIES IN EXPERIMENTAL SCIENCE. Illustrations, 15 × 23 cm. Cambridge, Harvard University, 1950. *Case 4. The Atomic-Molecular Theory*, by Leonard K. Nash. 115 pages. Price, $1.35 (paper bound).

When President Conant first offered his ideas on the understanding of science through an historical approach so that students of the humanities might relate developments in natural sciences to those in other fields of human activity, his platform did not offer him scope to treat the subject fully. Particular cases have been treated in the series of Harvard Case Histories. The fourth in the series treats with the pivotal stages in the development of the modern atomic and molecular theory.

The booklet deals in detail with the deceptively simple proposals advanced by John Dalton; the achievements made in relation to the then contemporary chemistry with the aid of these proposals; the discovery by Gay-Lussac that suggested a better substitute for Dalton's assumption; Avogadro's reconciliation of their apparently incompatible viewpoints; and a sketch of the steps which led to the acceptance of his views.

Few major advances in scientific knowledge have been made in such a remarkably short space of time (five years) and an uncanny intuition was required to formulate the steps in the advance of a doctrine which had vaguely existed in men's minds for two thousands of years. The whole account is treated by Dr. Nash in scholarly fashion and provides an excellent introduction to a subject which is exercising the minds of many non-scientific men today.

T. C.

MATHEMATICS, QUEEN AND SERVANT OF SCIENCE, by E. T. Bell. 437 pages, illustrations, 14 × 21 cm. New York, McGraw-Hill Book Co., 1951. Price, $5.00.

This book is a revision and enlargement of two former publications, "The Queen of the Sciences," and "The Handmaiden of the Sciences," by the same author, who has also written "The Magic of Numbers" and "Men of Mathematics." Professor of Mathematics at California Institute of Technology, E. T. Bell is well qualified to interpret and popularize mathematics for the non-mathematician. This is exactly what is done in "Mathematics, Queen and Servant of Science," except that a lot of history of mathematics has been included.

There is also much to be found in this book by the mathematician. There is a vast difference between the undergraduate problem-type of mathematics and the postgraduate pure mathematics. It is books such as this, which introduce the student to many of the concepts of pure mathematics while still not omitting the applied mathematics with which he may be familiar, that should be required reading for all those undergraduates majoring in mathematics. Familiarity with these concepts should enable the student to decide whether he wants to study pure mathematics and teach, or applied mathematics and perhaps be an industrial mathematician.

It may seem surprising that certain topics in postgraduate pure mathematics can be presented understandably to one having no background in mathematics, yet this is done.

DONALD B. HOUGHTON

PRINCIPLES OF PHASE EQUILIBRIA, by F. E. W. Wetmore and D. J. LeRoy. 200 pages, diagrams, 16 × 24 cm. New York, McGraw-Hill Book Co., 1951. Price, $3.50.

A clear understanding of the principles involved in phase equilibria is basic to the study of many of the physical sciences. This is particularly true in the fields of mineralogy, metallurgy and physical chemistry. In writing this book the authors have drawn upon their teaching experience and have produced a text in which the subject is developed in terms of fundamental principles and laboratory criteria. Mathematical theory and formal thermodynamics have been avoided in the body of the text.

The book consists of nine chapters and two appendices. The first chapter contains a general discussion of chemical equilibrium. Chapter 2 is devoted entirely to one-component systems. The next three chapters deal with systems consisting of two components. Chapters 7, 8 and 9 treat equilibria in three and four-component systems.

The first appendix contains a thermodynamic derivation of the phase rule. The second appendix shows curves of free energy *vs.* composition for systems under various conditions.

The book is printed on a good grade of paper and contains many diagrams, illustrating the material discussed in the text. There are exercises for the student at the end of each chapter which will give him an opportunity to apply his knowledge to the construction and analysis of phase diagrams.

As a text for a short course in phase equilibria or as a supplementary text in a regular course in metallurgy or physical chemistry, the book is to be highly recommended.

M. D. EARLE

LAPLACE TRANSFORMATION THEORY AND ENGINEERING APPLICATIONS, by William Tyrrell. 230 pages, illustrations, 15 × 22 cm. New York, Prentice-Hall, 1950. Price, $3.75.

In this book the author has presented the theory of the Laplace transformation in such a form as to be readily useful to the average engineer. At no time is the continuity of the development beclouded by excessively rigorous derivations, but at all times sufficient rigor is present so that the reader may proceed with confidence and understanding. The presentation gains from the author's inclusion of many physical problems, the solutions to which are well known by other mathematical methods. The resultant comparisons serve well to illustrate the advantages of the Laplace transformation as well as to give a physical meaning to the solutions. In addition, the author has included many problems for the reader to work out on his own. The uses of the Laplace transformation which are covered in this text are dynamical applications, structural applications, partial differential equations, difference equations and closed loop systems.

EZRA KRENDEL

BOOK NOTES

INVENTORIES OF APPARATUS AND MATERIALS FOR TEACHING SCIENCE. VOLUME II, UNIVERSITIES, 146 pages; VOLUME III, TECHNICAL COLLEGES, PART I, VETERINARY SCIENCES, 98 pages. 15 × 24 cm. Paris, United Nations Educational, Scientific and Cultural Organization, 1950. No price.

These two books are part of a series of three volumes published in several languages by UNESCO and compiled from reports submitted by experts in the field of science education from six countries considered leaders in scientific development. The series is designed to aid science teachers in planning curricula and selecting essential equipment for the better training of science students in under-developed and war-torn countries. Volume II covers botany, chemistry, geology, physics and zoology at the University level. Volume III, Part I deals with veterinary sciences, as a branch of the Technical College program. Courses of study are outlined, followed by lists of the necessary equipment to be used, with approximate prices given in U. S. currency. A comprehensive, well-done and very useful series.

DIRECTORY OF INTERNATIONAL SCIENTIFIC ORGANIZATIONS. 224 pages, 15 × 24 cm. Paris, United Nations Educational, Scientific and Cultural Organization, 1950. No price.

This directory gives information on international, non-governmental organizations whose activities deal with various branches of education, science, and culture. Under separate headings are listed the names of the officers and aims of each organization, financial resources, meetings, history, publications, etc. The directory is divided into three main sections: Basic Sciences, Applied Sciences and Miscellaneous. In each of the three sections, the entries are listed alphabetically, and a subject index further aids the reader to locate quickly the information he is seeking. Designed for specialists, the directory contains a wealth of material not heretofore published in compact form.

METHUEN'S MONOGRAPHS ON PHYSICAL SUBJECTS. Illustrations, 11 × 17 cm. London, Methuen & Co. Distributed in U. S. by John Wiley & Sons, Inc., N. Y.

Semi-Conductors, by D. A. Wright. 130 pages. 1950. Price, $1.75.

This volume presents material on the theory of electron flow in semi-conductors and across the boundary between them and either a metal or a vacuum. It is an elementary treatment of the theory of semi-conductors, intended for students of electron physics of solids. Practical applications and experimental results are treated but briefly, for the main emphasis is on the basic theory. Up-to-date references are provided at the end of each of the eight chapters, for those who wish to investigate the subject more thoroughly.

NOTES FROM THE BIOCHEMICAL RESEARCH FOUNDATION

ELLICE McDONALD, M.D., DIRECTOR

S^{35}-GLUTATHIONE PREPARATION FROM YEAST AND TRACER STUDIES IN CANCEROUS AND NON-CANCEROUS RATS

BY

GLADYS E. WOODWARD

Glutathione (GSH) occurs in relatively high amounts in almost all living tissue yet its exact function has not been established. The use of S^{35}-labeled GSH in tracer studies has not previously been reported. In the present studies, S^{35}-labeled GSH has been prepared and administered to normal, protein-depleted and tumor-bearing rats, and the S^{35} uptake in various tissues has been determined.

PREPARATION OF S^{35}-LABELED GSH

The synthesis of S^{35}-labeled GSH in quantity sufficient for tracer studies had not been described, although Franklin (1) showed that radioactive GSH was present in yeast grown in a medium containing S^{35} in the form of sulfate. In attempting to grow yeast on a scale sufficiently large to allow isolation of GSH, we found that certain cultures of yeast contained very little GSH but that the GSH concentration in yeast was regularly high if a low concentration of glucose and a high rate of aeration were maintained in the medium during growth of the yeast. Description of our most satisfactory procedure for production of radioactive GSH through biosynthesis by yeast is therefore given somewhat in detail.

The basic medium contained the following materials per liter of culture:

Salts—$NH_4H_2PO_4$, 6 gm.; KH_2PO_4, 2 gm.; $MgSO_4 \cdot 7H_2O$, 0.51 gm.; $MgCl_2 \cdot 6H_2O$, 1 gm.; KCl, 2.1 gm.; NaCl, 0.1 gm.; $CaCl_2 \cdot 2H_2O$, 0.13 gm.; $CuCl_2$, 0.1 mg.

Growth Factors—asparagine, 1.5 gm.; inositol, 10 mg.; calcium pantothenate, 4 mg.; thiamin HCl, 0.4 mg.; pyridoxin HCl, 0.4 mg.; nicotinic acid, 0.4 mg.; biotin, 20 μg.

Five hundred ml. of basic medium to be used for growth of the yeast inoculum were autoclaved 15 min. at 120° C. in a 2200-ml. Fernbach culture flask. For the main culture, salts and growth factors for 7 liters of medium were used. To the dissolved salts, in a 22-liter round bottom flask, were added 1 ml. of sodium lactate (60 per cent), $H_2S^{35}O_4$ solution[1] equivalent to 1–2 mc. S^{35}, and water to 6000 ml. This solution was auto-

[1] The S^{35}-labeled H_2SO_4, obtained from the Oak Ridge Laboratory of the U. S. Atomic Energy Commission, was carrier free and contained less than 0.02 mg. S per ml.

claved 45 min. at 120° C. The growth factors were combined in about 200 ml. of water. This solution was autoclaved separately for 15 min. at 120° C. and then added to the autoclaved solution of salts. The 22-liter flask was then fitted with a sterilized stainless steel aeration tube (disc diameter $2\frac{3}{8}$ in., porosity coarse, 65 microns), a dropping funnel for adding glucose and NH_4OH, a tube for withdrawing samples, and an exit tube for the air.

The 500 ml. medium was inoculated with a 48-hr. culture of *Saccharomyces cervisiæ* grown in 10 ml. of nutrient glucose broth. Five ml. of 50 per cent glucose were added and the culture was shaken gently on a horizontal shaker at room temperature during the growth of the yeast. Further additions of 5 ml. of glucose were made when the previous addition had been used up. Concentrated NH_4OH was added also when the pH of the culture was found to be below 4.0 in order to bring the pH up to 5.5–6.0 and to provide additional nitrogen for yeast growth. After about 48 hours of growth, when centrifugation of a small aliquot showed the culture contained about 15 ml. of packed yeast cells, the culture was added to the medium in the 22-liter flask. Sixty ml. of 50 per cent glucose were added to make the initial concentration of glucose about 0.5 per cent. Aeration was started and then maintained as rapid as possible without loss of material by foaming over. From time to time, glucose was added in such amount that the concentration in the medium was never greater than 0.5 per cent, and NH_4OH was added to maintain the pH of the culture in the range of 4–6. When about 500 gm. of glucose had been utilized, 60–72 hr., the yeast was centrifuged from the medium and washed by centrifugation. Data on yeast yield and GSH production are given in Table I.

The GSH in the moist packed yeast was isolated as described by Schroeder, Collier and Woodward (3). Additional GSH was recovered

TABLE I.—*GSH Preparation in Radioactive Yeast Cultures.*

Culture No.*	Sulfur Supplied, mg.	Glucose Supplied, gm.	Yeast Yield: Packed Volume, ml.	Yeast Yield: Dry Weight, gm.	GSH Content of Yeast** (Total), mg.	Total S Incorporation in: Yeast, %	Total S Incorporation in: GSH, %	GSH Isolated,** mg.
I	500	540	600	120	546	96	11	205
II	500	395	460	91	667	90	14	158
III	1000	1323	804	178	720	63	7.5	274

* I and II cultured as described in text; III cultured similarly except double amounts used throughout and glucose concentration of 1 per cent (maximum) used.
** GSH content of sample determined by glyoxalase method (2).

TABLE II.—*Analytical Data on Radioactive GSH Preparations.*

Preparation No.*	Total S Content Found (Theoretical = 10.42%), %	Total GSH Content, Found by Glyoxalase (2), %	S Content Calculated from GSH, %	Radioactivity (of 1 mg.), μc.
III	6.4	60	6.3	0.106
IV	7.3	69	7.2	0.157

* III from Culture III (Table I); IV from Cultures I and II combined before recrystallization.

as described by Hopkins (4). Analytical data for once-recrystallized samples are given in Table II. Although the two GSH samples were only 60 per cent and 69 per cent pure, the contaminant could not have been a sulfur compound since the total sulfur was entirely accounted for by the sulfur from GSH. Since the non-sulfur impurity derived from yeast would not be expected to interfere in animal experiments in which the utilization of the GSH was to be traced by means of the radioactive sulfur atom, and since further purification would have resulted in a lower yield of GSH, the GSH samples were used without further purification.

TREATMENT OF ANIMALS

Male albino rats raised on Purina Laboratory Chow were used. Rats with subcutaneously growing Walker carcinoma No. 256 about 2 cm. in diameter served as the tumor-bearing animals. Rats maintained on a low-protein diet for 15–18 days, as in the experiments of Leaf and Neuberger (5), served as the protein-depleted rats. The rats were given radioactive GSH as described below and killed by decapitation 18 hr. after administration of the single dose or of the last dose when more than one dose was given.

Rat 1. A tumor-bearing rat, weighing 130 gm., was given a total of 50 mg. (3.94 μc.) of radioactive GSH (Preparation III) forcibly by mouth over a period of 3 days in 5 portions freshly dissolved in 0.5 ml. of water.

Rat 2. A protein-depleted rat, weighing 82 gm. after being on the low-protein diet for 15 days, was given a single subcutaneous injection of 50 mg. (3.52 μc.) of radioactive GSH (Preparation III) freshly dissolved and neutralized in 1.6 ml. of water.

Rat 3. A protein-depleted rat, weighing 83 gm. after being on the low-protein diet for 18 days, was given a total of 100 mg. (6.84 μc.) of radioactive GSH (Preparation III) forcibly by mouth over a period of 3 days in 5 portions freshly dissolved in 0.5 ml. of water.

Normal and Tumor-bearing Rat Groups. Three rats weighing 200 ± 4 gm. were used in each group. They were fasted 30 hr. before dosage and during the assimilation period. To each rat 40 mg. of radioactive GSH (Preparation IV), freshly dissolved in 1 ml. of water, were fed by stomach tube in a single dose. The total radioactivity of the dose was 2.08 μc. for the tumor-bearing rats and 0.34 μc. for the normal rats. (The lower radioactivity in the latter case was due to radioactive decay during the time between experiments.)

TREATMENT OF TISSUES

The tissues were removed immediately after decapitation of the rats. Individual tissues from Rats 1, 2 and 3 were dried *in toto* in an oven at 100° C. and then powdered. Corresponding tissues from the three rats of the normal and tumor-bearing rat groups were pooled. The pooled

fresh tissues were extracted with trichloroacetic acid and the GSH was isolated as the cuprous mercaptide from the extracts as described by Anderson and Mosher (6). The residual protein, after 6 extractions with trichloroacetic acid, was washed twice with alcohol and twice with acetone and air-dried to a powder.

MEASUREMENT OF RADIOACTIVITY

Samples of tissue protein and GSH were prepared for radioactive sulfur assay essentially as described by Anderson and Mosher (6): $BaSO_4$ samples, prepared[2] from the powdered tissue preparations, were weighed into aluminum counting cups; analyzed aqueous solutions of GSH were evaporated in the counting cups. The amount of $BaSO_4$ or of GSH in the cups was usually less than 1 mg. per sq. cm. The assay of a given amount of S^{35} was found to be the same whether in the form of $BaSO_4$ or GSH.

The radioactivity of the samples was determined either in a "Q-gas" counter[3] or in an internal type counter designed in this laboratory.[4] Counts obtained in either counter could be correlated since both were calibrated with standard S^{35} samples.

Corrections in the count for self-absorption and resolving time, where necessary, were made from experimentally constructed curves. Sufficient counts were taken with individual samples to give a statistical accuracy of about 2 per cent. Check counts made with two or three widely different amounts of each preparation gave specific activity values which agreed usually within a few per cent.

RESULTS AND DISCUSSION

The first experiment, with a tumor-bearing rat, showed (Table III, Rat 1) that the tissue sulfur was replaced by sulfur from the administered

TABLE III.—*S^{35} Uptake in Tissues of Rats Administered Radioactive GSH.*

Rat No.	Determination	Tissue								
		Liver, %	Kidney, %	Spleen, %	Small Intestine (Whole), %	Muscle (Leg), %	Tumor, %	Blood, %	Lung, %	Bone Marrow, %
1	S replacement*	0.81	0.95	1.02	1.04	0.44	1.04			
	Total S content**	0.84	1.27	0.94	1.07	1.04	1.05			
2	S replacement*	5.4	6.4	4.2	6.2	0.7				
	Total S content**	0.64	0.96	0.91	0.91	0.95				
3	S replacement*	6.6	5.0	4.2	6.4	0.74		4.1	4.0	6.6
	Total S content**	0.56	1.11	0.68	1.08	1.13		0.92	0.78	0.55

* Per cent S replacement = $\frac{\text{specific activity of S isolated}}{\text{specific activity of S administered}} \times 100$. The values for per cent replacement may be converted to *Standard Specific Activity* (7) by a factor of 0.400 for Rat 1, 0.252 for Rat 2, and 0.127 for Rat 3.
**Based on dry weight.

[2] The $BaSO_4$ samples were prepared by members of the Microchemical Department of this Laboratory.

[3] Model D 46 from Nuclear Instrument and Chemical Corporation, Chicago.

[4] Five samples at a time could be placed within this counter tube. A standard sample was always included. After evacuating the tube, the quenching gas was provided by introducing ethyl alcohol vapor to a pressure of 2 cm. (Hg) and then argon to a total pressure of 16 cm. (Hg).

GSH to the extent of about 1 per cent in all of the tissues studied with the exception of muscle which took up somewhat less. Since no tissue of this rat, which was in a good state of nutrition, was outstandingly selective for the S^{35}, it was thought that selectivity might be shown in an animal in a poor state of nutrition. However, when protein depleted rats (Table III, Rats 2 and 3) were used, again no selectivity was observed, but the degree of S^{35} uptake in the tissues with the exception of muscle was 4 to 8 times greater than in the tumor-bearing rat while the total sulfur content of the tissues was not significantly different. The difference in S^{35} uptake appears to be due to the poor state of nutrition in the depleted rats rather than to the presence of a tumor in the adequately fed rat, since later experiments (Table IV) showed that essen-

TABLE IV.—*S^{35} Uptake in Tissue Protein and Tissue GSH of Cancerous and Normal Rats Administered Radioactive GSH.*

Tissue	Replacement* of Sulfur in:				Ratio of S Replacement	
	Protein		GSH		GSH/Protein	
	Tumor Rats, %	Normal Rats, %	Tumor Rats, %	Normal Rats, %	Tumor Rats	Normal Rats
Intestinal mucosa	2.200	1.904	—**	7.20		3.8
Tumor (non-necrotic)	0.780		2.98		3.8	
Liver	0.564	0.460	2.84	3.44	5.0	7.5
Blood	0.327	0.233	1.29	1.56	4.0	6.7
Kidney		0.528				
Spleen		0.299				

* Per cent S replacement = $\frac{\text{specific activity of S isolated}}{\text{specific activity of S administered}} \times 100$. The values for per cent replacement in these rats may be converted to *Standard Specific Activity* (7) by the factor 0.768.
** Experimental material lost.

tially the same amount of S^{35} was taken up by the tissues of rats with tumors as by tissues of non-cancerous rats under the same conditions of nutrition. The S^{35} uptake in the two protein-depleted rats was about the same in spite of the fact that the dose and mode of administration were different. It appears, therefore, that the degree of S^{35} uptake is determined to a much greater extent by the nutrition of the animal than by the dose or mode of administration of the GSH.

Direct comparison of the S^{35} uptake from GSH in cancerous and non-cancerous rats, under identical conditions of nutrition, weight, dosage and mode of administration, showed (Table IV) that the S^{35} was found in both the protein and GSH separated from the tissues of both groups of rats. The degree of S^{35} uptake was not significantly different in corresponding tissues of the tumor-bearing and normal rats, either in the protein fraction or in the GSH fraction. The concentration of S^{35} in tumor protein and tumor GSH was within the range of the other tissues studied.

In the tissue protein of both groups of rats, the pattern of sulfur distribution from the administered GSH was similar to that reported from methionine (7,8) and from cystine (6) after the same assimilation period. The labeled sulfur atom was taken up more selectively by far in intestinal mucosa than in any other tissue studied. The degree of uptake of S^{35} from GSH was somewhat less than reported from methionine but more than from DL-cystine.

In view of the observation of Waelsch and Rittenberg (9) that GSH rapidly undergoes degradation and resynthesis in tissues, it is unlikely that the radioactive GSH isolated from the tissues contained intact molecules of the GSH administered. It is interesting to note, however, the large amount of radioactive GSH, 7.2 per cent replacement, in the sample isolated from normal intestinal mucosa. The degree of radioactivity of sulfur in the GSH isolated from the different tissues of both normal and cancerous rats was 3.8 to 7.5 times greater than in the protein from the same tissue, and in general the tissue GSH followed the same order of selectivity as the tissue protein. Almost the same ratios had been observed (6) for the distribution of S^{35} between the GSH and protein sulfur of liver and intestinal mucosa when DL-cystine was administered to rats. These results from GSH, together with those from cystine, lend support to the suggestion of Waelsch and Rittenberg (9) that GSH may function as an intermediate in the transfer of amino acids to proteins.

SUMMARY

S^{35}-labeled glutathione has been synthesized biologically by yeast grown in a medium containing S^{35} in the form of sulfate.

S^{35} from the labeled glutathione administered to normal rats or to rats with Walker carcinoma No. 256 was found in both the protein and the glutathione of the tissues. The concentration of S^{35} in the sulfur of the glutathione isolated was several times greater than in the protein sulfur of the same tissue.

The degree of uptake of S^{35} was very nearly equal in like tissues of the normal and of the tumor-bearing rats. The degree of uptake in tumor was about the same as in liver but considerably less than in intestinal mucosa.

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CURRENT TOPICS

Research and Development in the Department of Defense.*—The research and development program of the Department of Defense, which averaged a half billion dollars in volume between 1945 and 1950, has now passed the billion dollar mark and is still growing. In fact, Congress is now appropriating as much for military research and development alone as it appropriated for both the War and Navy Departments together in 1939. A large portion of this money is going to universities and other non-profit organizations and to industrial laboratories engaged on a variety of projects under contract to military departments. The impact on the national research and development capacity is tremendous; together, the expanding programs of the Department of Defense and the Atomic Energy Commission will soon absorb something like three-fourths of the entire national research and development effort and a similar portion of scientific and technical manpower.

The question which naturally occurs to the layman is what kind of thinking and planning are going into the preparation of the military programs and are these enormous sums being spent to the best advantage? Military research and development is unquestionably better organized today than at any time in the Nation's history. Prior to World War II, both the War Department and the Navy Department had their own research and development programs which where quite independent of each other and uncoordinated. In 1941, the Office of Scientific Research and Development was established as an independent agency reporting directly to the President and was largely responsible for the brilliant scientific achievements of the United States in World War II. The OSRD was a purely wartime agency, however, and was liquidated as of December 31, 1947. To fill the gap which the OSRD left, the National Security Act of 1947 (Unification) established the Research and Development Board of the Department of Defense which formulates the over-all program of military research and development and coordinates the programs of the three military departments.

The Board is not an operating agency, it maintains no laboratories and lets no contracts. Its function is to work closely with the Secretary of Defense and with the Joint Chiefs of Staff in order that new weapon developments may receive full consideration in military planning and the formulation of strategy. It recommends to the Secretary of Defense the appropriate apportionment of effort among various important areas such as, guided missiles, atomic energy, aeronautics, biological warfare, human resources, ordnance, and so on through the many fields of modern science and technology that play an important role in warfare.

The seven-man Board which directs the activities of the Research and Development Board is made up of two high-ranking representatives from each of the military departments, one civilian and one military, and a civilian chairman appointed by the President. The present Chairman is Mr. William

* Prepared for this JOURNAL by Lee Anna Embrey, Staff Writer, Research and Development Board.

Webster who is on leave as Executive Vice-President of the New England Electric System. The Board staff is composed of representatives of the three military departments who work with civilian scientists and specialists in assessing and evaluating the military research and development effort.

Through the technical committees and panels of the Board, the civilian and military members of the staff, representatives of other federal agencies, and civilian experts whose services are loaned to the Government by private enterprise, have excellent opportunity to exchange information and points of view in their respective fields. Through regularly scheduled meetings and by means of special conferences and symposia, those who have knowledge of, and interest in the given subject, not only contribute their own special knowledge but are given the opportunity to learn what is being done elsewhere in the same field. This cross-exchange of information and the automatic coordination which results as a by-product are among the most useful contributions which the RDB makes to the military research and development program.

Each of the three military departments has its own vigorous program, focused principally in the needs of its particular Service. In a number of areas, however, it is possible for all three Services to receive the benefit of research being carried on by a single department. One of the initial tasks of the Board was to bring to light areas of overlap and unnecessary duplication.

Since the development of the crisis in Korea, the emphasis has shifted somewhat from long-range objectives to the rapid completion of problems admitting of early solution. The aim has been to produce as rapidly as possible weapons which would be of maximum benefit to the fighting forces. The provision of additional funds by Congress has speeded the process by making possible the construction simultaneously of several prototypes and the pursuit of more than one approach to especially difficult problems.

The chief limitation on the over-all effort is the immediate and the potential shortage in scientific manpower. The demands for technically trained manpower by industry, by the armed forces, by educational institutions, and other civilian institutions, require a careful consideration of the whole problem and there are groups, of which the Scientific and Manpower Advisory Committee of the National Security Relations Board (The Thomas Committee) is an outstanding example, that are offering plans and recommendations at the national level for the most efficient conservation of this important resource.

In summation it may be said that although the military research and development program is being ably directed and administered at the national level, the real measure of its achievements is to be found in the hundreds of industrial and academic research institutions throughout the country who are bringing to bear upon important problems of national defense, the best they can muster in brains, loyalty, and steadfastness of purpose.

"Shoran," Precision Bombing Aid, Now in Use.—"Shoran," the bomber pilot's distance-measuring "yardstick," which won distinction in the closing months of World War II as one of the most dramatic contributions of electronics to America's military strength, is being used in Korea, according to press dispatches from Fifth Air Force Headquarters.

Invented by Stuart W. Seeley, director of the Industry Service Laboratory of RCA Laboratories Division, and developed by scientists and engineers of

the Radio Corporation of America as an outgrowth of the study of television "ghosts," shoran makes it unnecessary for airmen to see the target in order to make a pin-point strike. The system was used with devastating effect in "blind" bombing over European battlefields, where it was credited with uncanny accuracy in spotting targets at distances up to 250 miles, and under any condition of visibility, night or day.

Shoran, like radar, employs the echo-timing principle in which distance is measured by the elapsed time between transmission of a radio-wave pulse and the receipt of its reflection. In the military application of shoran, the bomber transmits individual signals to two widely-separated ground stations whose locations in friendly territory are known with great accuracy. When the signals reach the ground stations, the pulses are retransmitted to the plane where special receivers and apparatus automatically compute the distance of the plane from each ground station. This information, together with the known distance between ground stations, determines the plane's location with respect to the target.

Although RCA began work on the shoran project in 1938, it was not adapted to military use until late in World War II, when development for the Army Signal Corps was completed. Shoran remained a closely-guarded military secret until 1946. First proof of its effectiveness in the last war came when shoran-aimed bombs demolished a 30-foot enemy bridge in Northern Italy which several runs by visual bombers had failed to destroy. Thereafter, it was widely used and with phenomenal results. Toward the end of the war in Europe, the amount of tactical bombing done in any area depended largely on the availability of shoran equipment.

Shoran bombing was employed not only against military targets, but with devastating effect against enemy personnel. On many occasions, rolling barrages of fragmentation bombs were laid down only a few hundred feet ahead of advancing Allied troops. Such barrages, delivered by planes completely hidden from the ground, had a tremendously demoralizing effect on opposing ground forces. Shoran is also well-suited for air support of landing operations, since planes equipped with the device can lay a barrage of aerial bombs precisely along a definite line, enabling air crews to release bombs at correct points for maximum impact along such a line.

During World War II, field tests made of shoran-equipped planes during photographic reconnaissance flights showed that the probable error often was not more than 50 feet, independent of altitude distance, and without the necessity of establishing control points (known absolute positions) in the area to be photographed.

Shoran equipment, including ground stations, can be readily transported by air and set up in a few hours. Equipment in the plane requires only one operator, who may be an existing crew-member such as navigator or bombardier. Each ground station also can be operated by one man, and two ground stations can furnish shoran service to a number of equipped planes within range.

Other possible military uses of the system include shoran navigation of remotely-controlled planes, and shoran position-indication for precise dropping of air-borne troops, weapons and supplies. In peacetime, the system has been used to map off-shore oil fields and to survey inaccessible jungle areas with incredible accuracy.

Fog-Proof Electric Camera (*Industrial Laboratories*, Vol. 2, No. 3).—Development of a new-type camera using invisible electrical images, which can produce pictures without being fogged even in active atomic radiation areas, has been announced by the Office of Technical Information, Department of Defense.

Dubbed "Two-Minute Minnie," the camera is the result of a year-and-a-half research program under sponsorship of the Signal Corps. Work leading to production . . . was undertaken by Haloid Company of Rochester, N. Y., and Battelle Memorial Institute, Columbus, Ohio. The model produces a finished 4 × 5 picture two minutes after the shutter is snapped. Its plates do not require chemical processing, and can be used repeatedly by merely wiping off the image.

Fundamental principle of the camera is the newly-discovered electrostatic, electrophotographic process. Light is recorded on a selenium-coated metal plate, previously sensitized by an electrical charge. Speed of the plate is approximately the same as ordinary orthochromatic black and white film. The camera uses an ordinary lens and shutter, and the same light sources as conventional cameras: sun or reflected light, photoflood, flash or ordinary bulbs.

Where light hits the charged plate, the electrical charge leaks off the sensitized material in proportion to intensity of the exposure, and is grounded on the plate. What is left is an invisible electric image.

When finely ground charcoal or anthracite coal powder is blown across the plate, it sticks; the more the electricity, the more powder remains. The powder image is transferred to ordinary paper or other material coated with an adhesive layer, and covered by a clear transparent film to protect the surface.

Tones of the pictures obtained are similar to the usual black and white photo.

New Instrument Tests Shuttle Tension.—The tension of yarns as they leave the shuttle during weaving is of importance to every textile mill. Faulty shuttle tension means defects in the filling. To practically eliminate filling defects, a new shuttle tester has been brought out by the Saxl Instrument Company of Harvard, Mass., well known for their work in Tensiometers and Tension Controls. The new test board gives a simple means to measure and adjust the tension while the yarn leaves the shuttle under an equivalent of performance condition.

In operation, a shuttle with quill inside is put on a stand. Yarn coming from it is threaded through the rollers of the tension meter. Then it is rewound on the axle of a special electric motor which has its stand on the right side. When the current is switched on, the yarn is pulled out of the shuttle and through the tension meter at the right speed.

If the filling tension is high, the tension meter shows a high deflection, say, 35 grams. If the tension is low, the meter gives a lower reading, say, 10 grams. Moving the adjusting means in the shuttle, the quiller or weaver can then make such adjustments as will produce the best tension uniformly for each shuttle.

Shuttle tension is an important factor particularly in the weaving of fine nylons, since their dye absorption, as well as physical appearances change after the filling is laid into the cloth with different tensions. From sensitive

synthetics to crepe yarns, differences in appearance in the finished fabric—called filling bars—can be practically eliminated if the uniform and right tension between individual shuttles is achieved.

The Saxl Shuttle Tester is precise and simple and can be operated just by plugging into a light outlet. The tensiometer mounted on the test board is removable, and can be used separately to check the tensions of any yarns whatsoever, be they in the warper, winder, quiller, slasher or other yarn preparatory machines.

Extracting Uranium from Phosphates.—The Atomic Energy Commission has developed a process whereby uranium in phosphates can be extracted economically in the production of highly refined fertilizer known as triple superphosphate.

In the production of triple superphosphate, phosphoric acid is produced at one stage. In the production chain, between the phosphoric acid stage and the final stage, it appears to be feasible with the new process to extract uranium from the phosphate by attaching special equipment to the facilities used in the production of the fertilizer.

It has been well known for some time that uranium occurs as a very minor component in phosphate deposits in the northwest and in Florida. Considerable research work has been done to develop a process whereby uranium can be extracted economically from these very low grade sources. It is hoped that the new process will open up this new source of uranium. The process is classified and cannot be described.

Jesse Johnson, Manager of AEC Raw Materials Operations, recently pointed out that in the production of phosphoric acid, sulfuric acid is used. For this reason, sulfuric acid available to industry for the production of fertilizer also assists the Commission's program for extracting uranium from phosphates as a byproduct of fertilizer production.

Photographic Methods of Uranium Determination.—A report describing several methods of photographic determination of the presence of uranium-bearing minerals in rocks and ores has been placed in open file. This work was completed as a part of the program on radioactive materials undertaken by the Geological Survey on behalf of the Atomic Energy Commission.

The photographic methods used by Survey Geologists to examine radioactive materials are described in detail. Paths traversed by charged particles (alpha particles) that have been emitted by uranium and its daughter products are recorded as black tracks in special photographic emulsions and are studied under a microscope.

This type of analysis gives information on the location and concentration of radioactive elements which can be obtained in no other way. The report specifically describes some new techniques in preparing different types of photographic emulsions for use in the study of radioactive rocks, minerals, and ores.

Entitled "The Preparation of Nuclear-Track Plates and Stripping Films for the Study of Radioactive Minerals," and prepared by L. R. Stieff and T. W. Stern, the report has been placed in open file for public inspection in the Geological Survey Library, Room 1033, General Services Administration Building, Washington, D. C.

Precision Platinum Thermometer for High Temperatures.—The Weiller Instrument Corp. announces a new precision Platinum Thermometer for high temperatures, consisting of the purest platinum wire obtainable embedded in a ceramic matrix. The latter does not react with platinum even near its melting point, and it maintains its insulating properties throughout. Precision measurements can be made to 1400° C., with the thermometer remaining operative to 1600° C. Supplied with four leads, the thermometer is mounted in a stainless steel tube ¼ in. in outside diameter and 14¼ in. long for temperatures up to 1000° C.; for higher temperatures it comes in a variety of ceramic tubes. The ice point resistance is 10 ohms.

A synthetic arm with simulated veins and a source of "blood" supply has been developed at the National Naval Medical Center for use in teaching medical personnel how best to give hypodermic injections and to withdraw blood from the body. "Skin" of the arm is flesh colored vinyl resin plastic, while veins and arteries are indicated by dark and light colored latex tubes. A colored liquid can be "pulsed" through the tubes to simulate blood flow. Development of other body sections, such as wrist, ankle, shoulder and neck is under consideration. The training device is inexpensive, all materials cost less than $15.

Skid-detecting devices now keep all B-47 and B-51 planes on the straight and narrow, no matter how hard pilots step on the brake pedals. The Ait Force plans to equip other planes with the devices, which have been used by railroads for some time and may have possible applications for trucks and busses. Of two distinct types, both devices "sense" a skid and release pressure on the brakes until the skidding condition stops, thus limiting skids to only a fraction of a second, reducing tire wear to a minimum and cutting landing roll to a theoretical near-minimum. The skid-detecting devices have been confined so far to use on bicycle-type landing gear, but tests are now being planned on tricycle-type gears.

A combination rifle and shotgun weighing 3½ pounds has been developed by the Army Ordnance Corps for the Air Materiel Command, which will use the new calibre .22/.410, M-6 in survival kits issued to Air Force flyers operating over Arctic and other uninhabited areas. The gun hinges at the stock to fold into a standard survival kit fastened to the parachute harness seat sling, carries ammunition in the stock. The rifle is accurate up to 200 yards, has good penetration power.

A radio-controlled lifeboat will chug up to some ditched airmen next year, guided by an operator in an Air Force SB-29 rescue plane which dropped the boat, and after picking up survivors sail away on a gyro-compass course for the nearest land, with sufficient fuel on board for an 800 mile trip. The 30-ft. A-3 rescue boat, dropped with a 100 foot parachute, will be powered by a four-cylinder engine in a watertight compartment. The chute will be jettisoned on landing. Radio controls will permit rescue plane to drop the boat's rudder board, open engine vents, crank the engine and steer the boar to the survivors. Subsequent movements can be controlled either by rescue planes or occupants of the boat. Transmitters for boat control have already been installed in SB-29s, complete system will be installed in all A-3 boats by early 1952.

Culture Dish Permits Four Simultaneous Tests (*The Laboratory*, Vol. 20, No. 3).—The "Blue Plate Special" at the corner restaurant usually arrives on a plate that has ridges separating the meat and vegetables from the potatoes and gravy. Saves side-dishes and dish-washing.

The same principle, with much more important consequences, has been utilized in designing a new type of culture plate for growing bacteria colonies and studying their reactions in different media.

Known as the Felsen Quadrant Culture Plate, the new development was devised by Dr. Joseph Felsen, Director of Laboratories and Research at the Bronx Hospital, New York City.

When Dr. Felsen first devised a divided culture plate it was intended for use in insolating intestinal pathogens in cases of infectious diarrhea. But other uses soon were found for a dish which accommodates four different media such as Endo's, SS, desoxycholate citrate agar and MacConkey's or EMB.

The dish consists of a 15 mm (inside depth) × 94 mm (O.D.) Petri dish which is divided into quarters by two smooth bisecting ridges 6½ mm in height. Standard Petri dish covers fit the Felsen Quadrant Culture Dish. The dish is made of annealed borosilicate glass to withstand steam sterilization and sudden temperature changes. The Felsen Quadrant Culture Plate can be used to demonstrate what, heretofore, four ordinary dishes were used to show. At the same time comparisons can be made quickly and conveniently. The new dish is useful in cases where the technician wishes to examine growth patterns on the same or different media—involving similar inoculations or inoculations of four different strengths of the same culture.

In addition the Culture Plate may be used for antibiotic assays, blood titer determinations, bacteriophage studies, etc. Where photographic records are maintained a picture of one plate can be substituted for pictures of four different plates. Educationally, the quadrant plate is a valuable teaching aid. (ED. NOTE: For further information, write to Fisher Scientific Company, 635 Greenwich, New York 14, N. Y.)

Melting Furnace (*Proceedings of the I. R. E.*, Vol. 39, No.4).—Development of a new, improved electric melting furnace has been announced by the Santa Anita Engineering Co., 2435 E. Colorado St., Pasadena 1, Calif.

This new unit has a built-in thermostatic heat control which automatically maintains a temperature within 20° F. of any given setting from 450° to 850° F. It is designed primarily for use with lead, lead alloys or other metals and materials with fusing point or melting point within its thermostatic limits of operation.

Materials are discharged into a mould or other receptacle through a spout located in the base of the pot. The spout valve is operated by a top handle lever, which remains cool during operation. By drawing the molten alloy off the bottom, each mould cavity is filled under the head of pressure, thus eliminating air pockets, resulting in full castings. Slag problems are also overcome by drawing off the bottom.

An ingot mould, which casts four ingots, is furnished with each furnace. The furnace is large enough to accept standard 5-pound (lead) ingots without cutting or prior melting.

New Weather-forecasting Techniques with G-E Analytical Mass Spectrometer (*Instruments*, Vol. 24, No. 3.—Scientists of the Geophysical Research Directorate, a branch of the Air Force Cambridge Research Laboratories, have announced that information being obtained by means of an instrument called the analytical mass spectrometer will be used in new studies of the composition of the earth's atmosphere. The studies, then explained, may lead to the development of new weather-forecasting techniques.

They emphasized that the high-sensitivity range of the mass spectrometer which was designed and built in the General Electric Company's General Engineering & Consulting Laboratory at Schenectady, N. Y., will aid in the study of reactions which meteorologists believe take place among constituents of the atmosphere as a result of absorption of radiant solar energy. It was pointed out that the studies are still in the preliminary stages.

G. E. engineers added that the mass spectrometer is so sensitive that under some conditions it can detect a gas which is only 1/100,000 part of another gas. Because it separates molecules of different masses, the instrument is useful in recording presence of isotopes, particles which react chemically in the same way but differ only in mass and atomic structure. Scientists of the Cambridge Air Force Laboratories will use the instrument to seek upper-atmosphere isotopes which have not yet been discovered.

Autos Counted by Robot that Sorts out Speeds (*Science News Letter*, Vol. 59, No. 14.)—A new vehicle-speed counter has been developed by two University of California at Los Angeles scientists which makes the task of checking highway traffic flow an easy one for the operator. He can relax while the machine does most of the work.

The new vehicle-speed counter can (1) tab the passing of a large number of cars, (2) segregate them according to speeds, and (3) record the speed of any given vehicle within an accuracy of plus or minus one thousandth of a second.

Daniel L. Gerlough and F. D. Schrieder of the University's Institute of Transportation and Traffic Engineering are the inventors of the device.

The highly rugged instrument, which can be handled by one technician, is about the size of a lady's vanity case and weighs only 33 pounds. A separate power unit is about the size of an automobile battery and weighs 34 pounds.

On the face of the instrument are ten counters which can be set to record various speed ranges: 0–15 m.p.h., 15–20 m.p.h., 20–25 m.p.h., etc.

Two detector strips are placed across the highway, three feet apart. When an automobile runs across the strips the machine automatically computes the difference in time between activation of the strips, translating this time difference into miles per hour.

Plastic boats turned out at the Puget Sound Naval Shipyard combine strength and durability with low upkeep cost. Present production model is a 12 foot wherry with maximum eight man capacity, empty weight of about 140 lbs. Injection-molded of liquid polystyrene, the boat features styrofoam for buoyancy instead of cork, eliminates caulking of seams and frequent painting required for wooden craft. The boats are now being turned out at the rate of one per day.